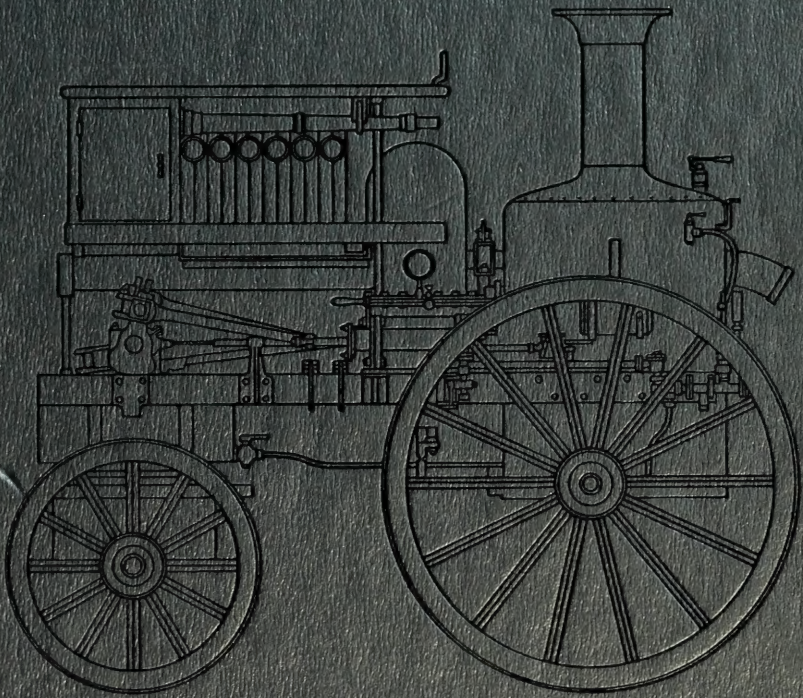

HOW THINGS WORK

HOW THINGS WORK

II

ILLUSTRATIONS RESEARCHED BY
ROGER JEAN SÉGALAT

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CONTENTS

HOW THINGS WORK

In the *piston type fountain pen* a screw spindle is connected to the filling cap and engages with a screw thread with which the hollow piston rod is internally provided. Attached to the front end of this rod is the piston which forms an air-tight and liquid-tight seal to the rear of the ink reservoir. When the filling cap is rotated, the screw spindle, rotating inside the piston rod, causes the latter to move longitudinally (Fig. 1a). When the filling cap is turned anti-clockwise (Fig. 1b), the piston travels forward. A small passage allows air to get behind the piston and thus prevents suction that would hinder the forward movement of the piston. When the piston is subsequently retracted, the air escapes by the same path. The return movement of the piston is produced by turning the filling cap clockwise (Fig. 1c). The suction that is thereby developed in the bore of the reservoir draws ink into the pen and causes the reservoir and the ink channels to be filled (up to the air hole of the nib). The pen must, during filling, be dipped in ink to above the level of the air hole.

When the pen is in use, the ink flows from the reservoir through capillary grooves in the feed; it thus reaches the underside of the nib and eventually finds its way along the slit to the tip of the nib. The pressure exerted by the writer causes the two points of the nib to be splayed farther apart, so that the slit widens and allows more ink to reach the tip and thus be transferred to the paper. In proportion as ink flows out of the reservoir, air must enter it. For this purpose the feed is provided with an air passage through which tiny bubbles of air make their way into the reservoir (Fig. 1d).

The ink cartridge of the *cartridge type fountain pen* (Fig. 2) is made of flexible plastic and sealed by a glass ball. It contains about 1 cm^3 of ink. When the cartridge is inserted into the pen, the ball is thrust back into the cartridge by a pin and thus releases the flow path of the ink. The function of the feed is to conduct the ink from the cartridge to the nib (Fig. 2b). At the end of the feed is a tongue which extracts the ink from the cartridge. Besides, the feed is so designed as to allow air to enter the cartridge. An internal sealing cap prevents ink escaping sideways from the mouth of the cartridge and seeping into the barrel of the pen. The "thermic regulator" consists of two tubes, one within the other, in the feed. The space between these tubes is so dimensioned that the ink clings to the walls by capillary action. The object of this system is to accommodate the excess ink which may flow out of the cartridge as a result of a rise in temperature (caused by body temperature in holding the pen) and which, in the absence of any compensation system, would cause a blot on the paper. The compensation "chambers" (i.e., the spaces between the tubes) absorb and hold the ink by their capillary action and thus provide a safeguard against blots.

Fig. 1 PISTON-TYPE FOUNTAIN PEN

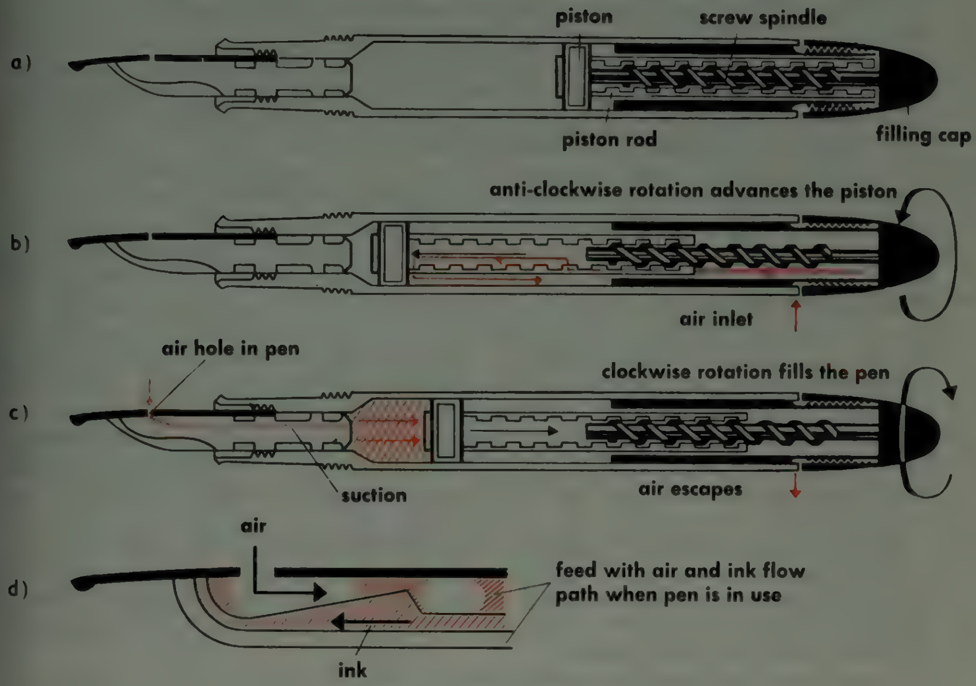
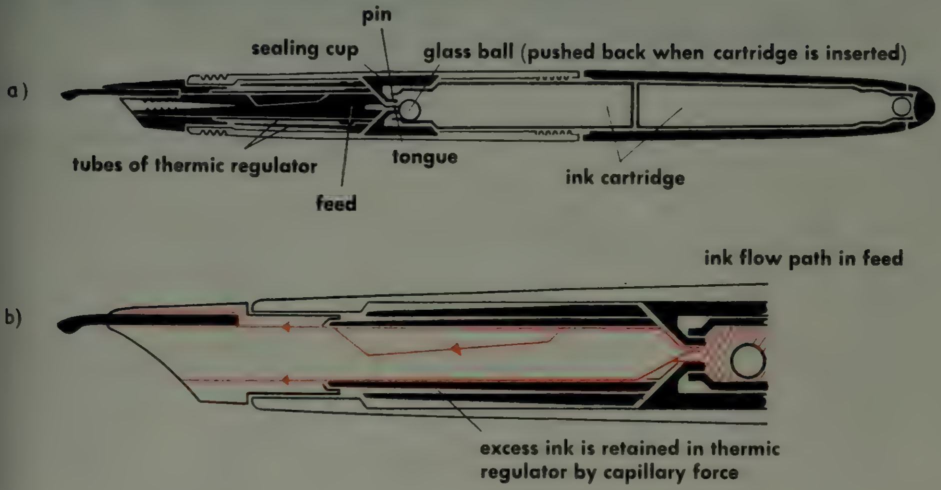


Fig. 2 CARTRIDGE-TYPE FOUNTAIN PEN



BALLPOINT PEN

The ballpoint is a writing instrument in which a small rolling ball, housed in a socket at the tip, transfers a viscous ink on to the writing surface. The ball is lubricated by the ink, which moves downwards in the reservoir chiefly by the action of gravity. When the pen is not in use and the ball is therefore not moving, it seals the end of the reservoir and thus prevents the ink from drying out. The ink is a viscous liquid containing either an oil-soluble dye or a spirit-soluble dye. The ball is usually made of steel, but in some pens synthetic sapphire balls have been used. Most manufacturers use a ball 1 mm (0.04 in.) in diameter.

In some ballpoint pens the reservoir and the tip comprising the ball can be retracted into the body of the pen when the pen is not in use. There are various mechanical contrivances for retracting and extending the tip of the pen. In one type of mechanism (Fig. 1), when the push-button is pressed, it pushes forward the thrust tube together with the ink reservoir. When the push-button is pressed, the catches of the thrust tube are fully inserted into the fixed slots with which the top part of the body of the pen is provided (Fig. 2a). When the button is released, the action of the large spring retracts the reservoir. The catches of the thrust tube (connected to the reservoir) engage with the small teeth on the rotating sleeve. The reservoir is then in the writing position (Fig. 2b). When the push-button is pressed again, the catches of the thrust tube plunge into the fixed slots (Fig. 2c). The rotating sleeve, which is spring-loaded by a small spring bears on the sharp edges of the fixed slots, while at the same time the sleeve rotates an amount corresponding to one tooth.

When the thrust tube moves back, the rotating sleeve is first lifted and, at the same time, turned. The catches of the thrust tube can then plunge into the large tooth gaps of the sleeve, so that the reservoir is fully retracted (Fig. 2d). The action is again controlled by the sleeve, which performs a small rotational movement whenever the push-button is actuated.

A much simpler form of retraction mechanism is illustrated in Fig. 3. It embodies a ball catch. When the push-button is pressed, the steel ball rotates in the clockwise direction in a heart-shaped cam recess the side of a cylindrical sleeve attached to the push-button (Fig. 4b). The position of the ball within the cam recess determines the position of the ink reservoir (Fig. 4c). When the push-button has been pressed, the ball is at the top holding point of the recess. It is held there by the pressure of the spring. The reservoir is then in the writing position. When the button is pressed again, the ball goes to the bottom holding point of the recess, and the reservoir slides back into the body of the pen.

Fig. 1 BALLPOINT PEN WITH ROTATING-SLEEVE PUSH-BUTTON ACTION

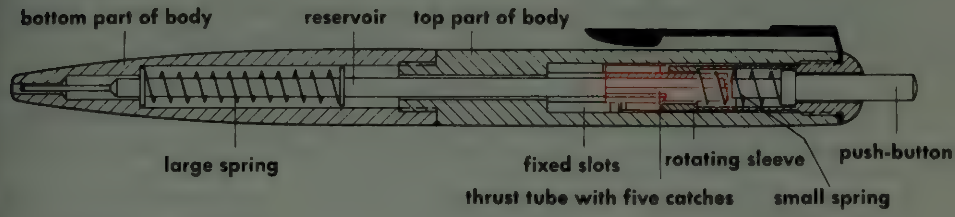


Fig. 2 OPERATING PRINCIPLE

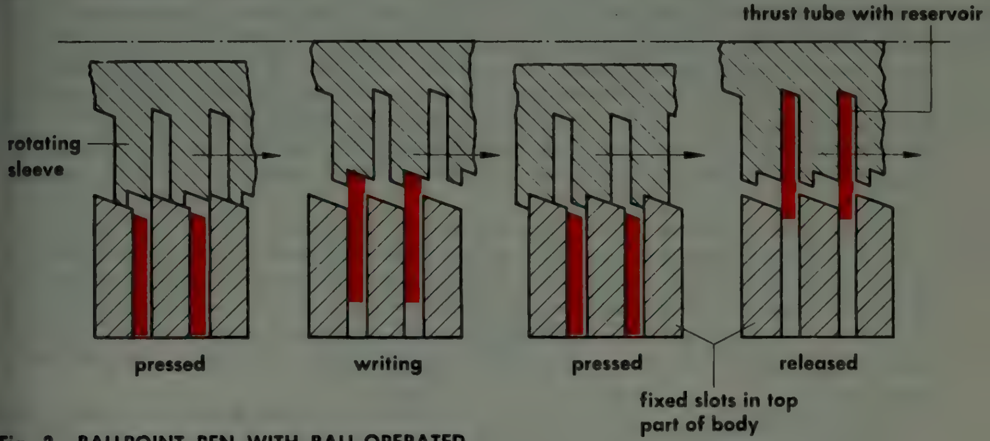


Fig. 3 BALLPOINT PEN WITH BALL-OPERATED PUSH-BUTTON ACTION

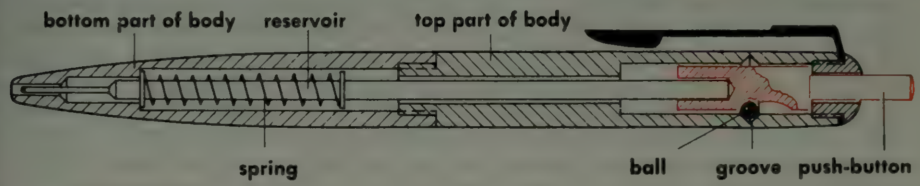
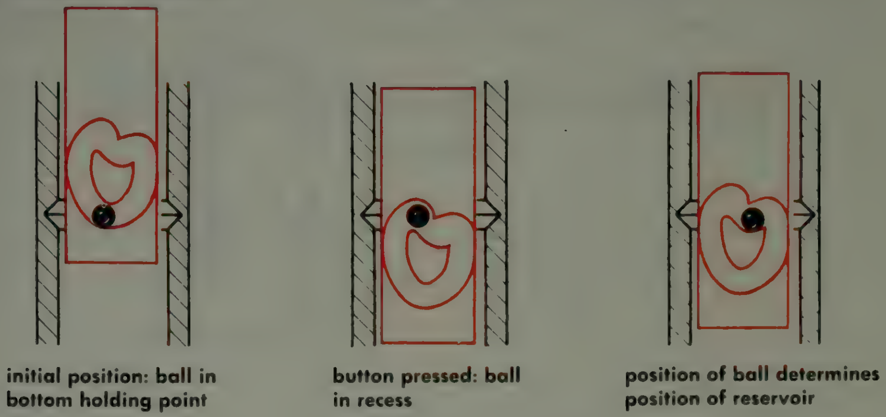


Fig. 4 OPERATING PRINCIPLE



The record player (gramophone or phonograph) reproduces the sound waves from the grooves in the gramophone record (see page 18). The vibrations set up in the stylus of the pickup are converted into electric signals. To make these sufficiently strong to work a loudspeaker, they have to be amplified and, if necessary, corrected by means of equalisers and tone controls. The record player (pickup and turntable), amplifier and loudspeaker constitute a reproduction system (Fig. 1). They are frequently combined into one unit, but for purposes of high-fidelity reproduction they are usually kept separate from one another, as they are otherwise liable to have an adverse effect on the quality of one another's performance (heat, vibrations, reaction).

A record player comprises a turntable with drive mechanism and a pickup mounted on a tone arm.

The drive mechanism (Fig. 2) must rotate the record (or disc) as noiselessly and free from vibration as possible at the correct speed. Vibrations in the running of the mechanism, causing irregular running, manifest themselves as objectionable "rumble" in the reproduction of the sound. To obviate this, the drive motor is resiliently mounted, and its motion is transmitted to the turntable through vibrator-damping drive components, such as a rubber friction gear. The latter usually also serves as a change-speed device, the change being effected by shifting the rubber friction disc longitudinally in relation to the motor shaft which is "stepped" so as to vary its diameter.

"Studio quality" drive mechanisms are usually provided with a heavy, accurately balanced turntable whose mass helps to equalise and compensate for any irregularities in the drive. Fluctuations in the speed of rotation manifest themselves as "wow" or "flutter" when they exceed about 3 per mille of the speed. The turntable should be made of a non-magnetic material (e.g., pressure die-cast zinc) in order to avoid adversely affecting the performance of the magnetic pickup (the type most frequently employed).

The function of the tone arm (Fig. 3) is to support the pickup and counterbalance its weight, so that the only forces acting upon the stylus are the deflecting forces exerted by the grooves. Any other force results in deterioration of the quality of reproduction. The "angular error" (Fig. 5a) cannot be entirely eliminated. When the grooves are being cut, the recording cutter moves radially, whereas the tone arm of a record player swivels about a pivot, so that its end, carrying the pickup, describes a circular arc. This difference between the direction of movement of the cutter and of the stylus causes distortion of the sound. However, by suitably shaping the tone arm (as in Fig. 5b), the angular error can be kept within acceptably low values throughout the path of the stylus on the record. Ideally, the tone arm should have frictionless motion. In high-quality arms the friction is virtually eliminated by means of precision ball bearings or jewelled bearings.

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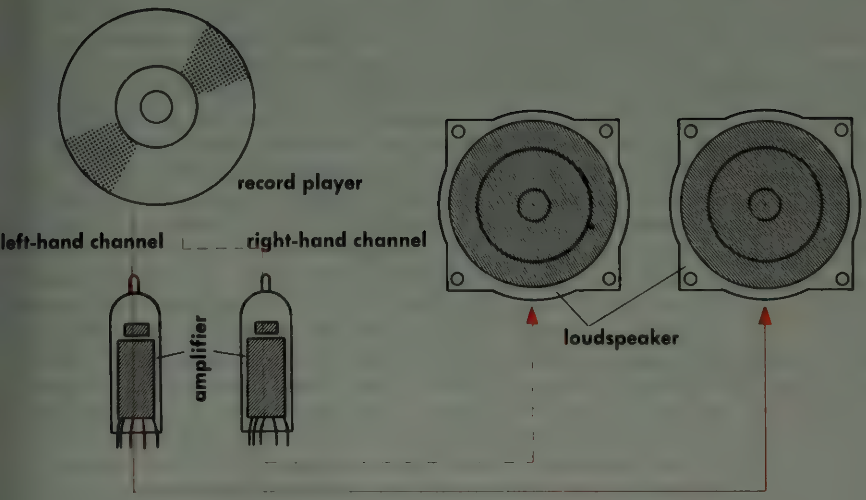
Fig. 1 PRINCIPLE OF STEREOPHONIC REPRODUCTION (*two channels*)



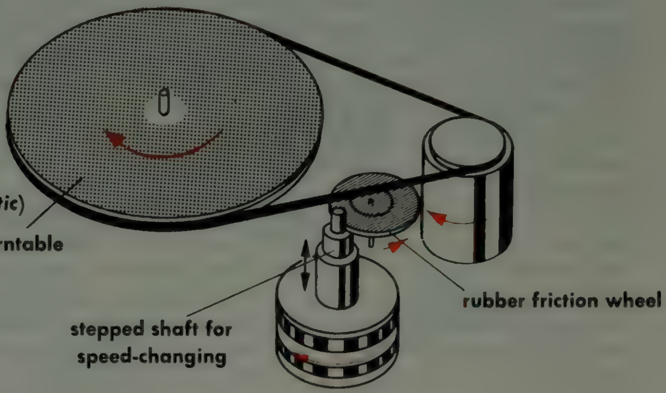


Fig. 2
DRIVE MECHANISM (*schematic*)

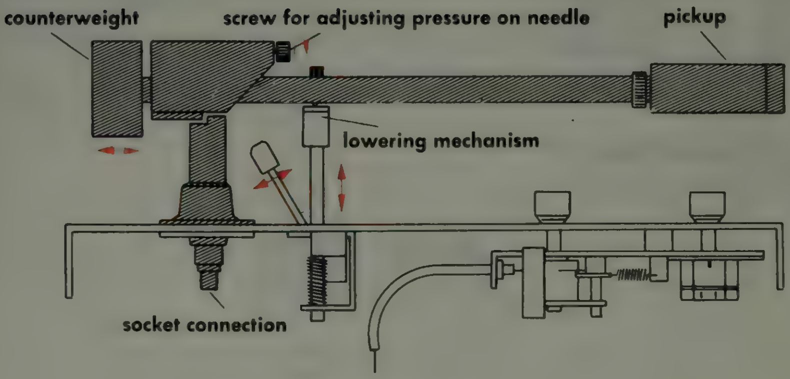


Fig. 3 SECTION THROUGH TONE ARM

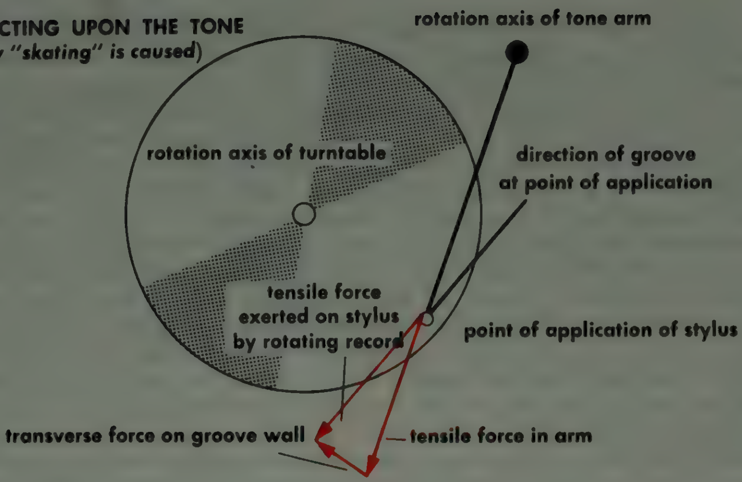
The effect of the inertia forces due to up-and-down movements of the tone arm can never be entirely eliminated. There is a limit to the reduction of the mass of the arm, as the latter must retain a degree of rigidity in relation to the modulations of the groove, so as to enable them effectively to move the stylus in relation to the pickup. A well-designed tone arm is carefully balanced. With regard to the vertical direction of movement this is achieved by means of an adjustable counterweight. Horizontal equilibrium is usually pre-adjusted by the design (mass distribution) of the arm and cannot subsequently be varied. The necessary vertical force to keep the needle in proper contact with the groove is produced either by a spring or by a weight.

Modern high-performance tone arms sometimes have a further refinement, namely, compensation for "skating", which is caused by the fact that the stylus presses harder against the inner than against the outer side wall of the groove. This is because of the transverse force component (Fig. 4). The compensating system must exert a counteracting (i.e., outward) force of the same magnitude upon the arm.

A tone arm must not resonate with the vibrations of the stylus, but should display rigid behaviour over the entire range of reproduction. Some tone arms are made of wood, which, because of its cellular structure, has particularly high vibrational rigidity. Steel tone arms of torsionally rigid tubular cross-section are widely used.

The pickup converts the movements of the stylus—imparted to it by the groove—into corresponding electric signals in the form of an alternating voltage. In a crystal pickup (Fig. 6) the deflections of the stylus cause a thin plate of Seignette salt (sodium potassium tartrate) to undergo flexural movements, which give rise to an electric potential difference in consequence of the piezoelectric effect (p. 104, vol. I). In a magnetic pickup (Fig. 7) the electric output is due to the relative motion of a magnetic field and a coil (or two coils in the case of a stereophonic pickup) located in the field. Magnetic systems are generally more complex and more expensive than crystal systems, but produce less distortion and are therefore almost exclusively used for high-fidelity reproduction. The stylus must have a rounded tip suited to the cross-sectional dimensions of the groove, so as to ensure that the tip is maintained in contact with the sloping side walls of the groove and clears the bottom (Fig. 8). The movable system in the pickup must be so mounted that the stylus can follow every rapid change of direction of the groove virtually without resistance; and for the same reason the mass of all the moving parts must be as small as possible. Pickups with very resilient stylus mountings can be operated with a relatively small vertical force on the stylus. Modern pickups combining high resilience with small moving mass so greatly reduce distortion and groove wear that deterioration of records due to repeated use has been virtually eliminated.

ig. 4 FORCES ACTING UPON THE TONE ARM (how "skating" is caused)



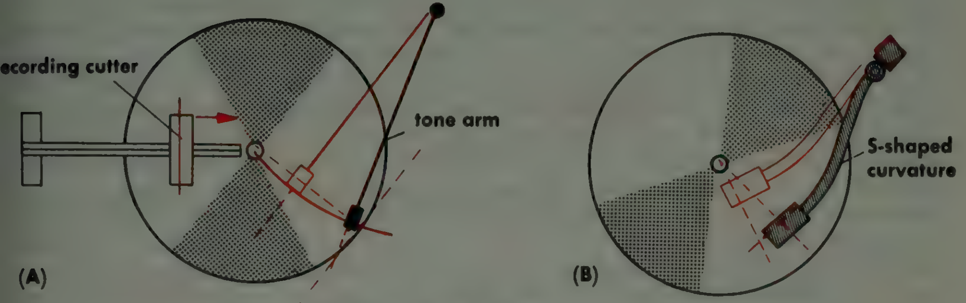


Fig. 5 ANGULAR ERROR (A) AND HOW IT CAN BE ELIMINATED (B)

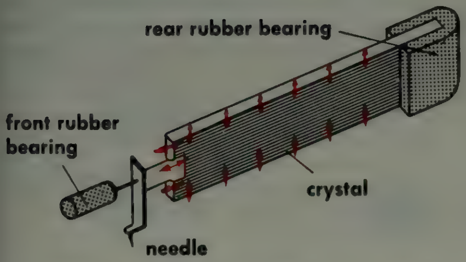


Fig. 6 CRYSTAL PICKUP, (*schematic*)

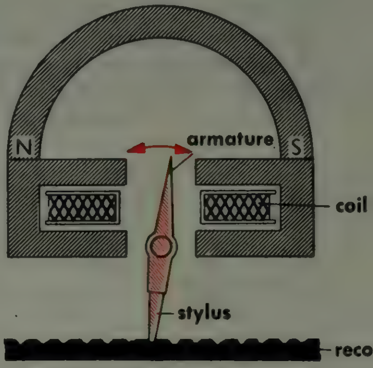
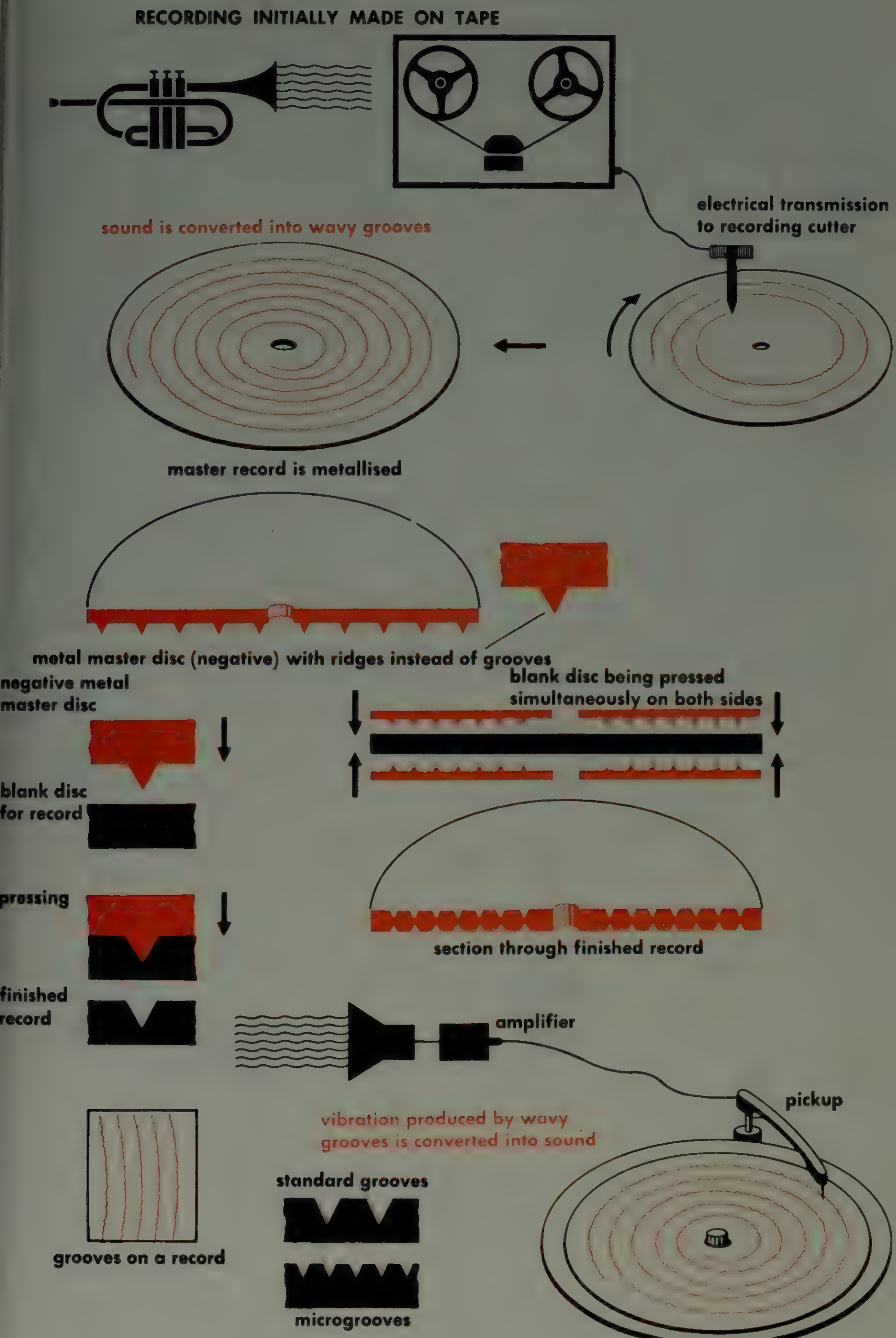


Fig. 7 MAGNETIC PICKUP (*schematic*)

Fig. 8 COMPARISON OF THE THREE GROOVE SIZES

A gramophone record (or disc) is a circular plastic disc on which the sound is recorded in a spiral groove. The sound vibrations of the air can be transmitted to a sharp recording cutter which scratches a wavy groove into a plastic-coated metal disc, the cutter being guided in a spiral path from the edge to the centre of the disc. The wavy pattern of the groove determines the frequency and the amplitude of the sound vibrations; these correspond respectively to the pitch and the loudness of the sound. On playing back this disc in a gramophone (at the same speed as the recording speed), the wave pattern of the groove will produce corresponding vibrations in the reproducing stylus, and these mechanical vibrations can be either directly (by means of a diaphragm, as in the early phonographs and gramophones) or indirectly (through a process of electrical amplification) be reproduced as audible vibrations of the air. In present-day recording techniques the master record is, of course, not itself used for playing back. The plastic surface of the disc is given a lacquer coating which is subsequently stripped off again after being metallised (plated with metal) by an electrochemical process known as electroforming. In this way an electroformed metal master disc is obtained. This disc is a "negative" in that the grooves in the master record are here represented by ridges. The negative can be used in a plastic moulding press to produce records. For commercial production of records in very large numbers, however, more durable negatives made of high-grade steel may be produced. These are then used as pressing masters. Modern fine-groove records rotate at $33\frac{1}{3}$ r.p.m. (long-playing records) or at 45 r.p.m. (extended-play records). The older standard records rotated at 78 r.p.m. Records for rotation at $16\frac{2}{3}$ r.p.m. are also produced for special purposes. The material of which the records are made plays an important part in determining reproduction quality. The cheaper records are made of synthetic thermoplastic resins containing fillers. With the introduction of fine-grooved records it became necessary to use high-quality compounds without fillers, as the filler particles produce a certain amount of background noise. A fairly wide range of synthetic materials are used for gramophone records; the vinyl plastics are most commonly used for the purpose, e.g., polyvinyl chloride. These are so-called "unbreakable" records. The moulding of records is done under pressure, with the aid of heat.

Stereophonic records are recorded with the aid of two separate microphones. Reproduction is accomplished by means of a single stylus with its axes of movement inclined at 45° to the record surface, the pickup being designed to produce two independent signals in accordance with the motion components along the two (90° displaced) axes. The two signals are amplified and fed to two separate loudspeakers.


RECORDING INITIALLY MADE ON TAPE
electrical transmission to recording cutter
sound is converted into wavy grooves
master record is metallised
metal master disc (negative) with ridges instead of grooves
negative metal master disc
blank disc for record
pressing
finished record
blank disc being pressed simultaneously on both sides
section through finished record
amplifier
pickup
vibration produced by wavy grooves is converted into sound
grooves on a record
standard grooves
microgrooves

Many radio broadcasts, all repeat broadcasts, and nearly all automatically repeated telephonic information (e.g., the "speaking clock" time service) are recorded on magnetic tape. To produce a tape recording, the sound (speech or music) is converted into corresponding electric signals by means of a microphone in conjunction with amplifier equipment. These electric signals (voltage oscillations) produce variations in the strength of a magnetic field. The signals are thereby recorded on a magnetic tape which is magnetised along its length in accordance with the signals impressed on it. In early tape recorders steel wire or tape was used, but nowadays a plastic tape provided with a coating of powdered red iron oxide (gamma-Fe_2O_3) is most often used as a magnetic recording medium. Red iron oxide particles, needle-shaped and about 1 micron in length, are widely used. The black oxide of iron (Fe_3O_4) is sometimes used for the same purpose. The oxide particles, which are applied to the tape in a coating mixed with a binder substance are strongly magnetisable and retain their magnetic properties almost indefinitely.

Sound is transmitted as pressure waves in the air. The lowest musical notes have a frequency of about 30 cycles per second; the highest notes of musical significance are about 4000 cycles per second (the highest audible frequencies are in the 12,000–16,000 cycles/sec. range). The "tone colour" (timbre), however, consists of a complex mixture of frequencies, due to harmonics which may have as much as six times as high a frequency as that of the fundamental tone of the sound. All these vibrations are picked up by the microphone, amplified and converted into variations in the magnetic field of an electromagnet in the recording head (Fig. 1), whereby these variations are recorded on the magnetic tape. To reproduce the sounds, the tape is passed over a similar head, called the reproducing head, at the same speed as that used in recording. The magnetism stored up in the tape induces voltage oscillations in the electromagnet coil of the head, and the electric signals thus produced are amplified and then used to energise a loudspeaker (Fig. 2).

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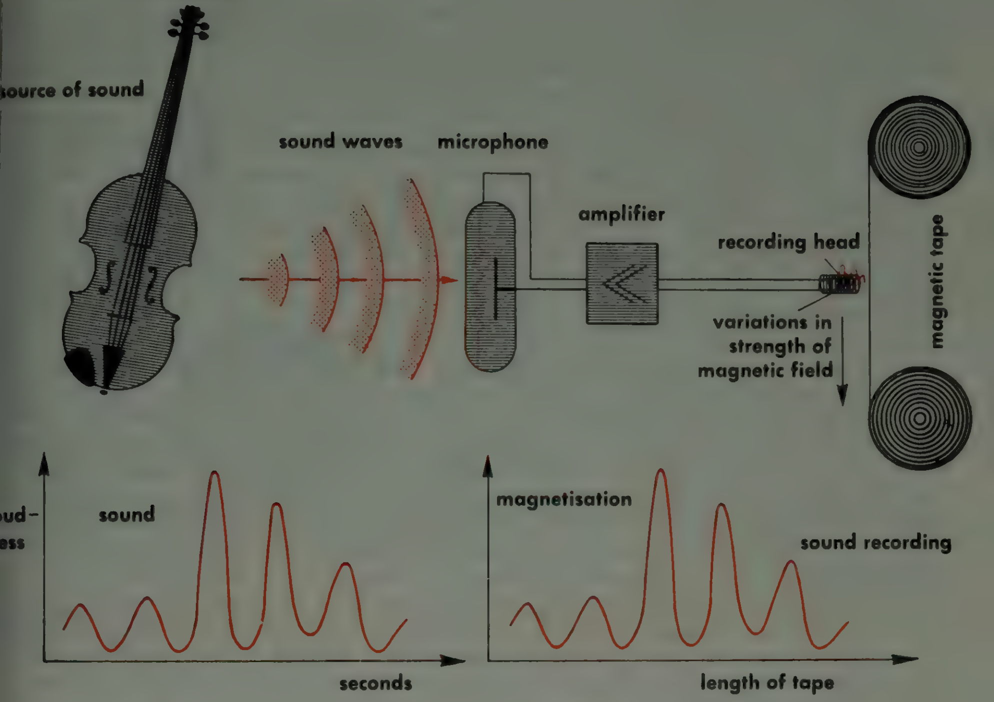
Fig. 1 RECORDING



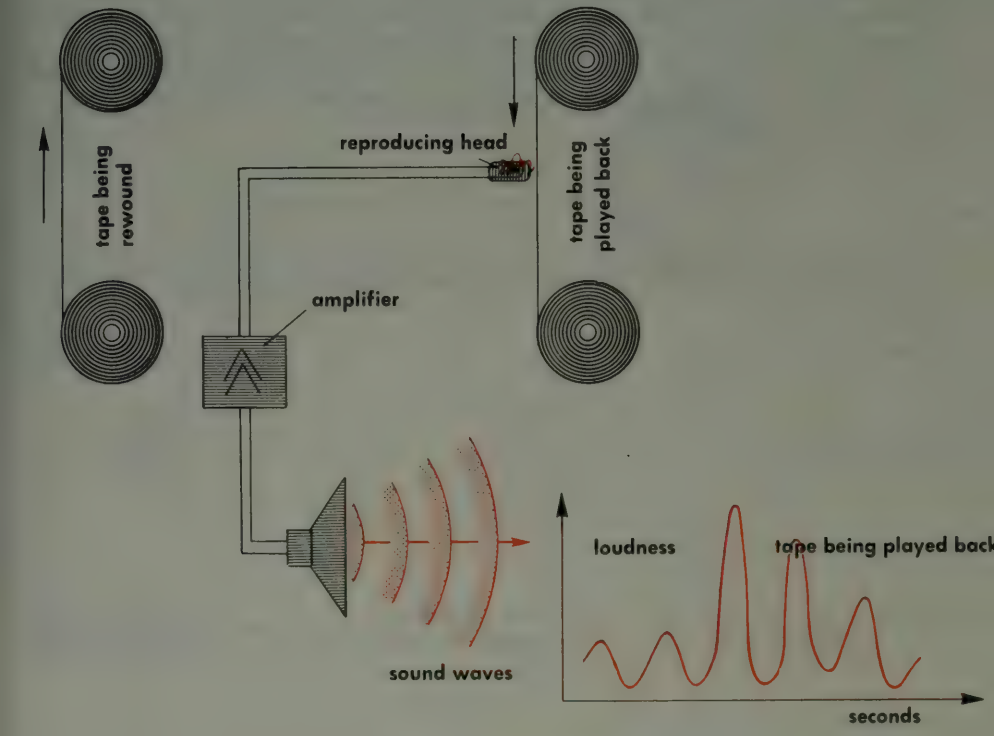


Fig. 2 REPRODUCTION

The recording head consists of a coil wound around a core of magnetic iron which has a gap at the point where the tape moves across its surface. The current in the coil magnetises the particles in the tape (Fig. 3). During playback, the whole process is reversed; indeed the recording head can also be used as the reproducing head (Fig. 4). The faster the tape travels past the recording head and, subsequently, past the reproducing head, the more faithful will be reproduction of the sound. This is because a higher tape speed provides more space for accommodating the highest frequencies on the tape. In order to record an overtone of, for example, 5000 cycles/sec. it is necessary to record 5000 oscillations in the strength of the magnetic field on the tape during each second of its passage past the recording head. In radio broadcasting studios a tape speed of 15 in./sec. is usually employed, i.e., each of the 5000 oscillations has a space of 15/5000 in. = 0.003 in. Tape recorders for amateur use are usually operated at tape recording and playing speeds of $7\frac{1}{2}$ in., $3\frac{3}{4}$ in. or $1\frac{7}{8}$ in. per second, many machines being provided with facilities for using any of these three speeds, as desired. The widths of tape available for the recording of each oscillation of the overtone of 5000 cycles/sec. are thus 0.0015 in., 0.00075 in. and 0.000375 in. respectively. There is a progressive decline in recording and reproduction quality as the speed is lower, since the width of the gap in the recording and/or reproduction head cannot be reduced indefinitely; each oscillation requires a certain minimum amount of space, which is greater according as the head (and more particularly the gap) is of coarser construction. The greater the fineness and precision of the head is, the more expensive is the tape recording equipment concerned. To achieve high-fidelity recording and reproduction, it is necessary to take a large number of technical factors into account. Sound recorders for professional purposes usually have three motors: one on the supply reel, one on the take-up reel, and a third the drive the tape. Extreme care is taken to provide smooth and uniform drive. Tape recorders for amateur use record usually on one-half of the $\frac{1}{4}$ in. wide tape. Some machines are designed to record four tracks on a $\frac{1}{4}$ in. tape. A great advantage of magnetic recording is that recordings can be erased and tapes re-used over and over again. Erasure is done by an erasing head, which produces a powerful high-alternating field that demagnetises and thus erases the tape just before the latter passes the recording head.

Fig. 3 RECORDING

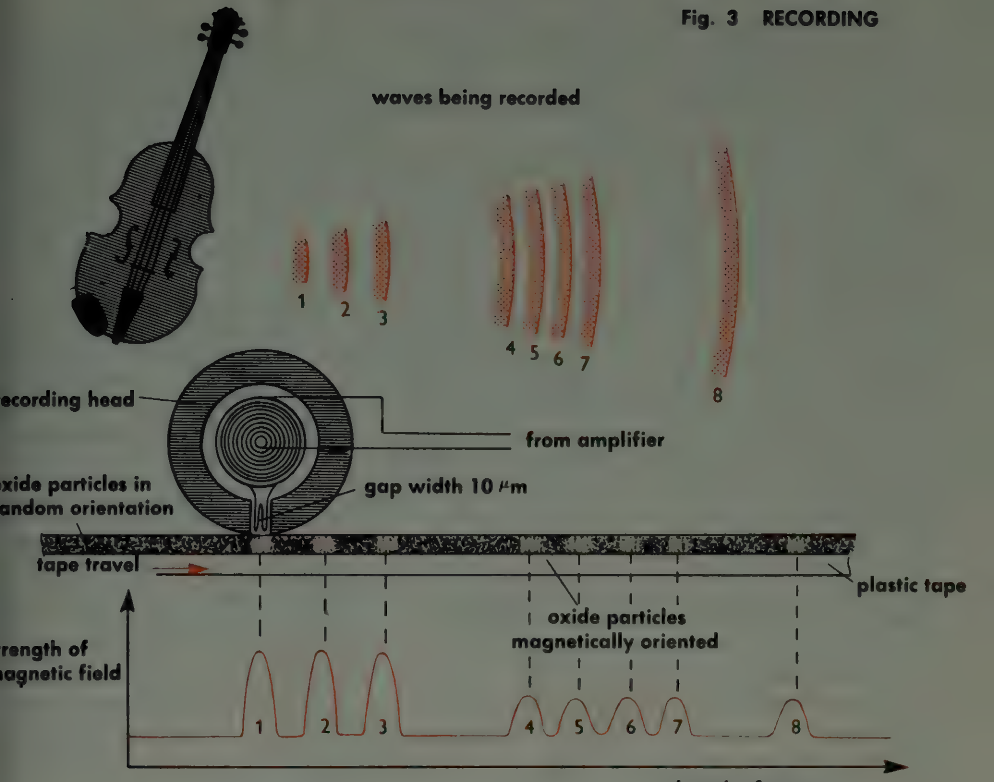
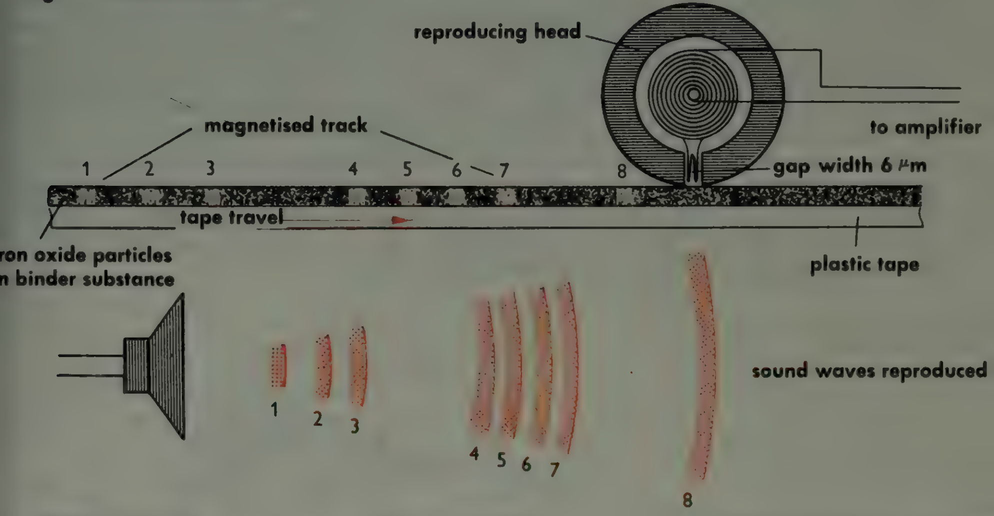
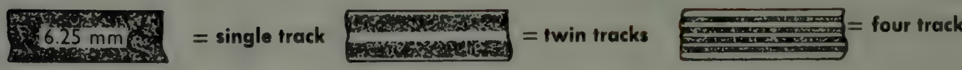


Fig. 4 REPRODUCTION





track widths

COIN TESTERS

A feature common to all "slot machines" (automatic vending machines, etc.) is the coin testing unit which rejects counterfeit or defective coins before they can release the locking mechanism. Every coin tester checks at least the diameter and thickness of the coin inserted into the slot; further tests relate to the weight, alloy composition, and magnetic properties. The coin inserted into the slot—which is just wide enough to admit the right coin—rolls through a chute and lands in the weighing device, which also serves to check the diameter of the coin: the two hook-shaped bent-round ends of the two balance arms are so spaced that a coin of the correct diameter will press down on the right-hand arm. A coin deficient in diameter will drop down between the hooks and into the receptacle for returned coins (vertical black arrows). Oversize coins are retained by the hooks, but as their centre of gravity coincides with the pivot of the balance, the latter will then not rotate. When the coin return button is pressed, the side flap of the coin tester is swung aside, so that the oversize coin is tipped over sideways and likewise falls into the returned coins receptacle. The counterweight on the left-hand arm of the balance ensures that only a coin having at least the same weight as this counterweight can pass. The "correct" coin now presses down the right-hand balance arm and rolls down a short incline, during which journey it passes through the field of a strong permanent magnet. Depending on the alloy of which it consists, the coin will be allowed to pass or will be retarded by this magnetic field. If the iron content is high, the coin may be stopped altogether. In that case it can be retrieved by pressing the coin return button, whereby a wiper dislodges the coin from the magnet, while the side flap is swung aside to allow the coin to drop into the return receptacle (black arrows on right). When the coin has passed the magnet, it drops on to a rebound stop. If the coin is "correct", it will have the right mass (depending on the alloy) and the right speed (depending on the degree of slowing-down by the magnetic field) and thus have the right amount kinetic energy to jump over the rejector pin and run down into the outlet giving access to the release mechanism of the machine (red arrows). Coins which fall short of this requirement will strike the pin and drop into the return receptacle. A somewhat different testing mechanism is used for coins made of plated steel (Fig. 2), as are used in some countries. The coin rolls along an inclined path whose slope is such that only a coin with a sufficient amount of kinetic energy will continue to move, whereas a coin that is too light will come to a standstill. Next, the coin is weighed in a device which functions like a trap-door: if the coin is too heavy, it will fall through the flap (right-hand black arrow). In the following stages the diameter and thickness of the coin are tested. The whole coin testing device is inclined sideways, so that, at a certain point, coins deficient in diameter automatically fall out sideways through an aperature between lateral guide plates. The thickness test is performed by letting the coil roll along a narrow ridge beside a gap. If the coin is too thin, it will drop into the gap (left-hand black arrow). Coins that are too thick are rejected by the slot in the first place and cannot enter the machine at all.

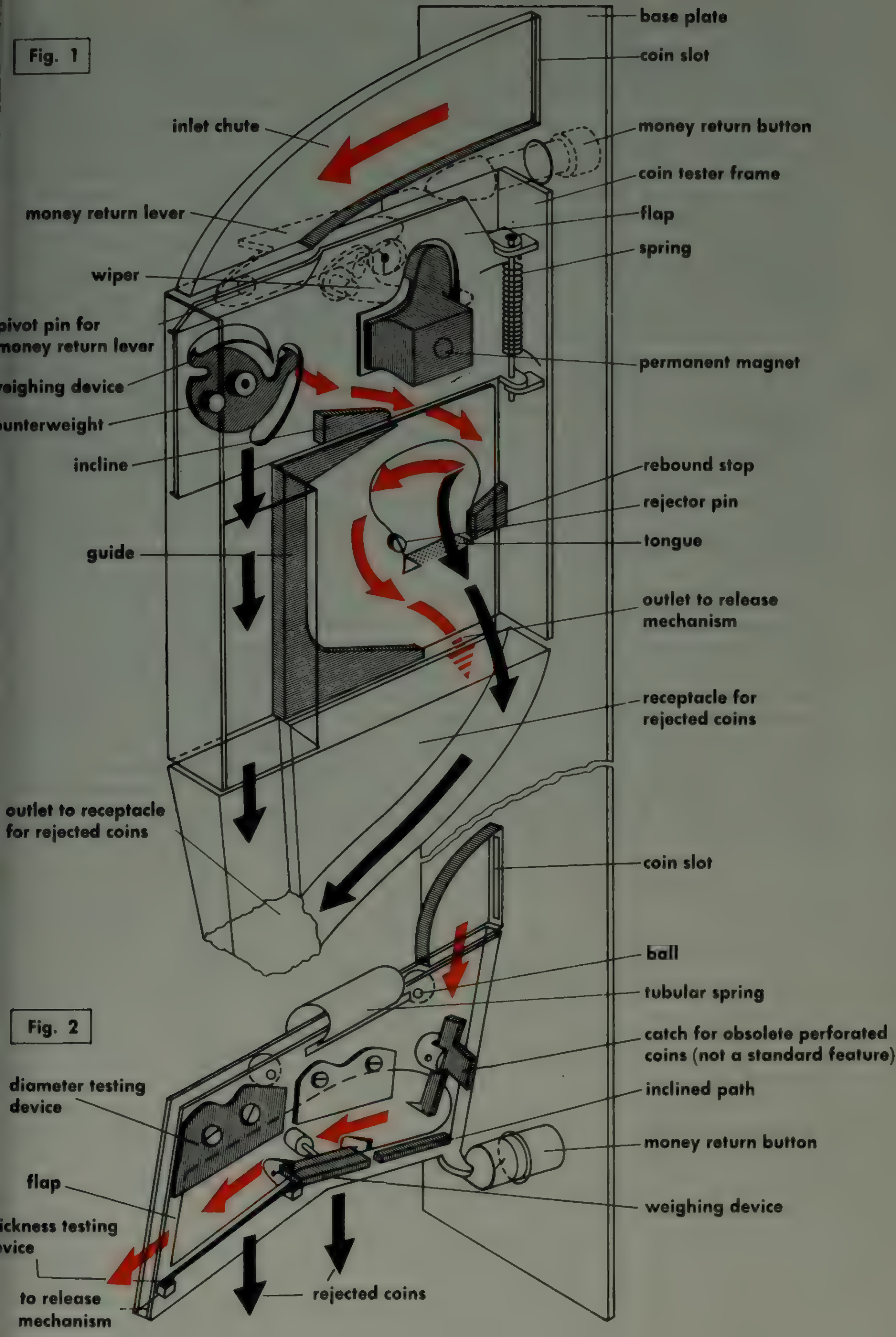
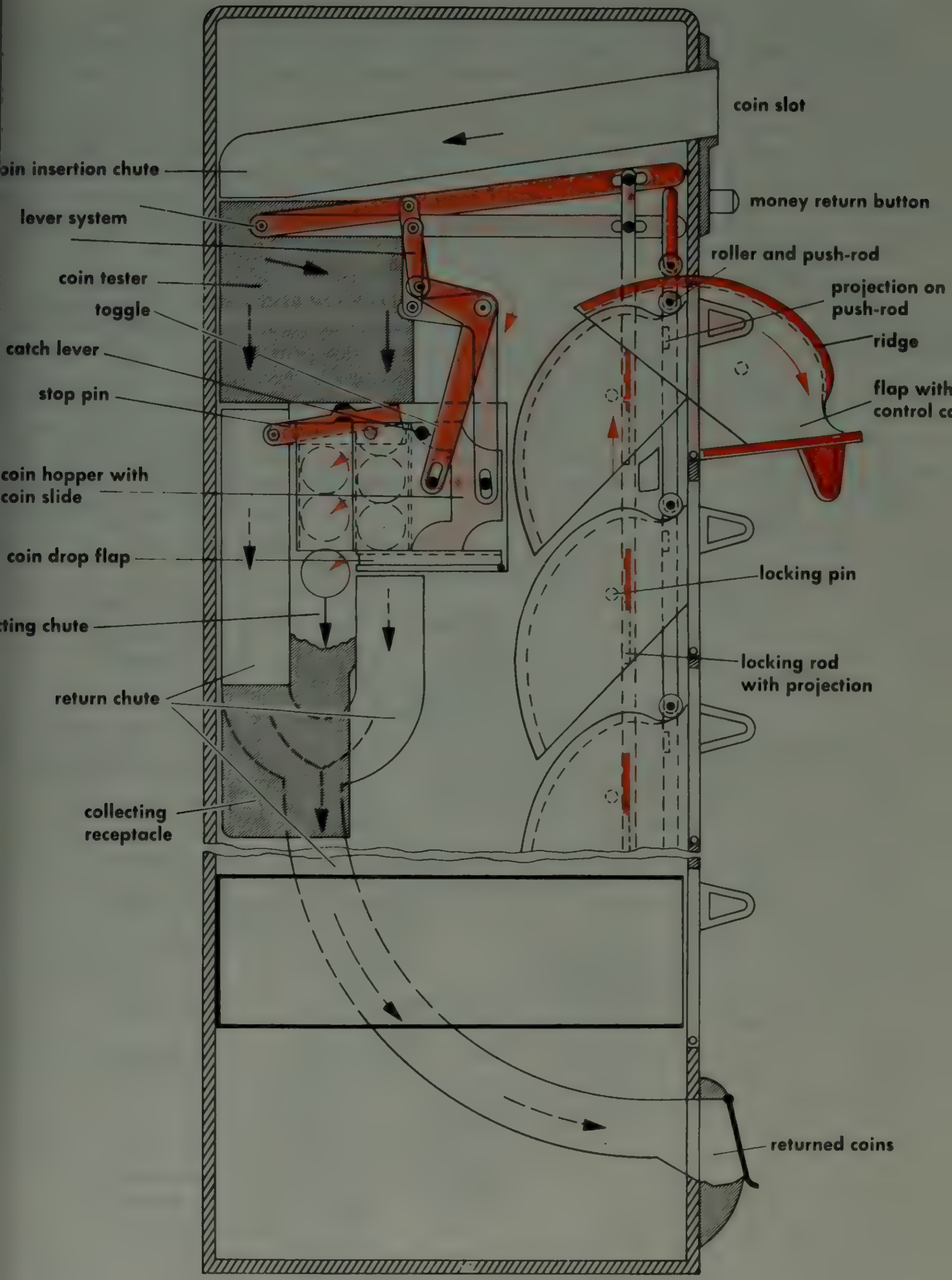


Fig. 1
base plate
coin slot
inlet chute
money return button
coin tester frame
money return lever
flap
spring
wiper
pivot pin for
money return lever
permanent magnet
eighing device
unterweight
incline
rebound stop
rejector pin
tongue
guide
outlet to release
mechanism
receptacle for
rejected coins
outlet to receptacle
for rejected coins
coin slot
ball
tubular spring
Fig. 2
catch for obsolete perforated
coins (not a standard feature)
diameter testing
device
inclined path
money return button
flap
weighing device
ickness testing
vice
to release
mechanism
rejected coins

These machines are usually equipped with a number of compartments which contain the goods for sale and which can be opened after the appropriate coins have been inserted into the slot of the machine. The operating principle of a typical machine of this kind is illustrated in the accompanying diagram. Each compartment is closed by a cover, or flap, provided with a glass window through which the goods can be viewed. Connected to each flap is a lateral control cam (curved path) which is in contact with a roller. After insertion of the appropriate coin or coins (in the machine illustrated it is assumed that three coins of equal value are required to work the machine) the purchaser pulls one of the flaps. This causes the cam to lift the roller and the vertical push-rod to which the latter is attached. Through levers this upward movement is transmitted to a toggle (elbow lever) which has a slot engaging with a pin on the coin slide, whereby the latter is given in a horizontal movement (to the left in the illustration). The top coin in the slide lifts the catch lever of a tongue projecting into the coin path. The coin slide can thus travel its full distance, allowing the flap of the compartment to be opened to its full extent and at the same time causing the coins to drop into the collecting chute. If on the other hand, no (or not enough) coins have been inserted, the coin slide travels only a very short distance, when a stop pin on the slide encounters the front of the catch lever. This locks the entire mechanism and prevents the flap from being opened. Any coin or coins inserted can, in that case, be retrieved by pressing the coin return button, which causes the bottom plate under the coin slide to swing down and allow the coins to drop in the return chute. In addition to this operating mechanism, a vending machine has to be equipped with various devices to prevent misuse, more particularly in the form of attempts to open more than one flap at a time after insertion of the money. This is prevented by a ridge at the edge of each control cam. When one flap is opened, the ridge on its cam slides under a projection on the push-rod under it, thus locking the push-rod and preventing the other flaps from being opened. After a flap has been opened, an additional locking rod frustrates any further attempts to open other flaps. This locking rod is connected to the top horizontal lever and extends down to the bottom flap. At the level of each locking pin (approximately at the centre of each control cam) the locking rod is provided with a projection which, when the rod is lifted, locks all the other—as yet closed—flaps. Cigarette vending machines operate on the same principle, except that, instead of compartments, there are slides or drawers which have to be pulled out and which are automatically refilled from a stack of packets.



coin slot
oin insertion chute
money return button
lever system
roller and push-rod
coin tester
projection on
push-rod
toggle
ridge
catch lever
flap with
control c
stop pin
coin hopper with
coin slide
coin drop flap
locking pin
ting chute
locking rod
with projection
return chute
collecting
receptacle
returned coins

Present-day juke boxes almost invariably use 7-inch records played at 45 r.p.m. Magnetic tape recordings and other recording media have not proved popular for the purpose. Most juke boxes hold between 30 and 100 records, which can be played on both sides.

The following components are common to all machines of this type (Fig. 1): record magazine or hopper, playing unit, programme panel, selecting device (usually provided with keys or buttons), coin tester and coin storage unit, amplifier, and loudspeakers. These components are built into a case with a transparent cover. Some juke boxes are additionally provided with a counting mechanism for the various records played (so-called popularity meter) and a remote-action selector (wall box) for remote control of the machine.

There are many design variants. The operation of a juke box will here be described with reference to a typical machine as illustrated in Fig. 1. After passing through the coin tester (see page 24), which may be equipped for testing two or three different values of coins, the coins drop into the corresponding coin hopper and actuate a micro-switch, which transmits an electric pulse to energise a magnet in the coin storage unit. This magnet in turn actuates a device which rotates the so-called adding cog-wheel, which has the same number of cogs as there are records in the juke box, e.g., 60. For example, the juke box may be so designed that it will play one record

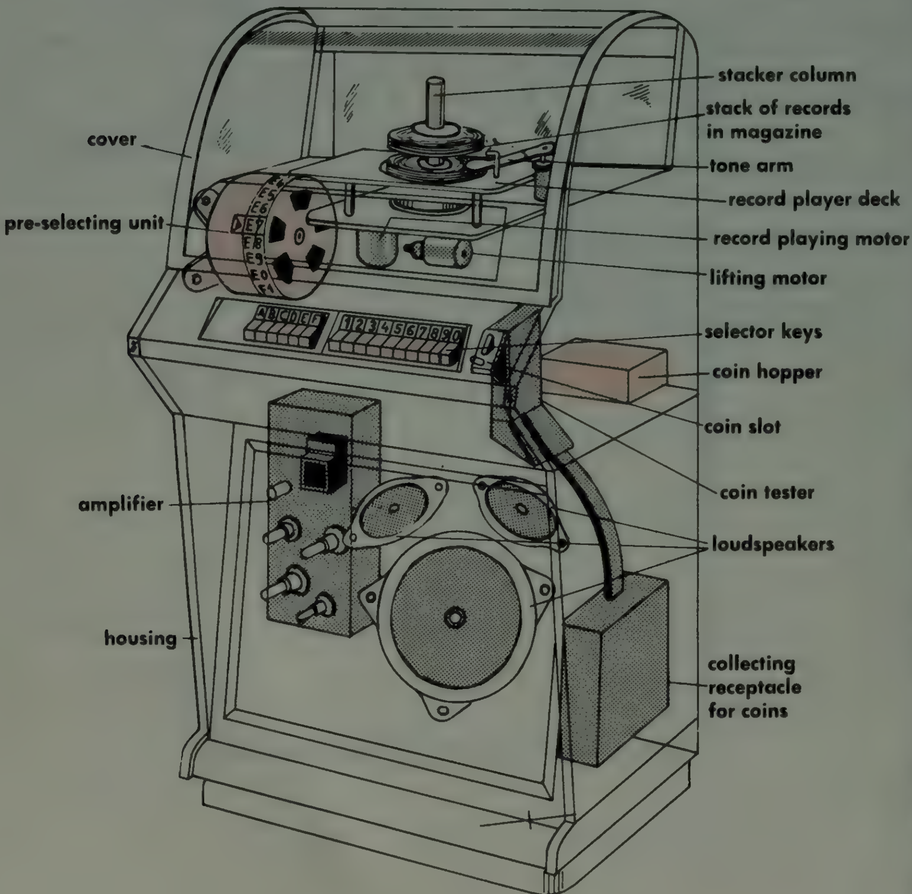


Fig. 1

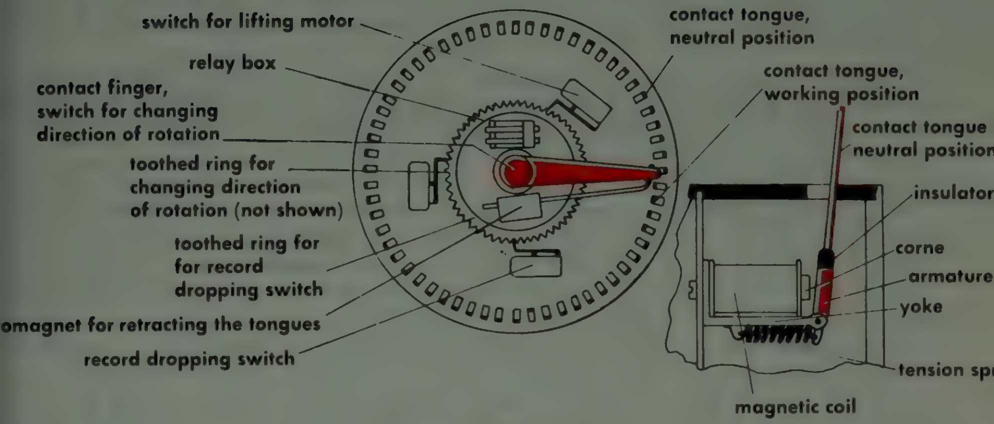


Fig. 2 PRE-SELECTING UNIT (*schematic*)

for sixpence, two for a shilling and five for a florin. (In the United States, one for a dime, three for a quarter.) On insertion of these coins the adding wheel will be rotated a distance corresponding to one cog, two cogs and five cogs respectively.

As a result a contact--situated between the adding cog-wheel and a second cog-wheel (called the subtracting cog-wheel)--is closed, so that current flows to the two unlocking magnets which then release the locking device that normally (i.e., so long as no coins have been inserted) prevents the selector keys from being operated. When the appropriate coins have been inserted, the desired record can be selected by pressing a letter key and a numeral key respectively (e.g., ten numeral keys numbered from 1 to 9 and 0, and six letter keys A to F, so that 60 combinations are possible). This causes micro-switches to energise a magnet, which causes the subtracting cog-wheel to rotate a distance corresponding to one cog. In this way the successive cog movements of the adding disc are cancelled one by one until the inserted money has been used up and the subtracting cog-wheel has caught up with the adding cog-wheel. When this happens the energising current of the unlocking magnets is cut-off and the keys become locked again. Simultaneously with the above-mentioned operations in the coin storage unit, an electric pulse travels from the actuated selector keys to the so-called pre-selecting unit, where it is stored.

The pre-selecting unit (Fig. 2) is the "brain" of the machine. It ensures that the pulse emitted by any particular selector key is transmitted to the desired record. This is accomplished as follows. An electromagnet associated with the selected key combination (there are, in this case, 60 such magnets) actuates a contact "tongue" (there are likewise 60 of these) which connects two contacts and thus completes an electric circuit to the selector motor (in the pre-selecting unit) and to a brake connected in parallel with this motor.

The selector motor drives a so-called relay box which has a contact finger that scans the contact tongues at the perimeter of the pre-selecting unit. When it encounters a tongue which has been pulled out by an electromagnet, the selector motor circuit is interrupted, and the motor is immediately stopped with the aid of the brake.

Until then a micro-switch which is connected to a toothed ring around the relay box produces a number of electric pulses corresponding to the number of teeth on the toothed ring (which is the same as the number of non-protruding contact tongues). These pulses are then fed to a dropping magnet inside the stacker column (Fig. 3). This magnet retracts three catches and, at the same time, causes three retaining tongues to protrude. Each pulse arriving at the magnet causes a gramophone record to be dropped down from the stack until the selected contact tongue has been reached. The stacked records actually rest on a carrier disc (Fig. 4), which descends a certain distance, according to the number of records dropped. See also Fig. 5.

Since each contact tongue corresponds to a particular piece of music, the "right" record has now been reached. Whether this record, too, is dropped down will depend on whether the protruding contact tongue corresponds to the "front" (top) or "back" (underside) of the record. In the selector key system (in this particular type of machine) an odd number corresponds to "back" and an even number to "front". For example, if the piece of music chosen is characterised by the combination B1, then, starting from the record A1/2 (which is at the bottom of the stack) the records A1/2, A3/4, A5/6, A7/8 and A9/0 will have to be dropped. Since B1 is the back (underside) of the sixth record, it is not dropped (the contact tongue B1 is protruding, no dropping pulse is transmitted): the swivelling tone arm, which has a pick-up provided with two stylus systems, plays the underside of this record from underneath. If B2 is chosen instead, i.e., the piece of music on the front (top) of this record, then the contact finger moves past the contact tongue B1; the corresponding tooth on the toothed ring of the relay box causes the micro-switch to emit a pulse; the dropping magnet is energised; the record B1/2 is dropped. Now the pick-up plays the front (top) of this record from above.

The swivelling motion of the tone arm is produced by the record playing motor (Fig. 5), which drives the turntable and (through rubber rollers) the stacker column and furthermore drives the mechanism for swivelling and lifting the tone arm. These various movements are performed by appropriate reversals in the direction of rotation of the motor. For example, when the turntable rotates clockwise, the tone arm is swung inwards and the pickup stylus applied to the record. When the record has been played, the direction of the motor is reversed, so that the turntable then rotates anti-clockwise and tensions a spring which lifts the stylus off the record and swings the tone arm outwards. At the same time, a lifting mechanism (comprising a motor, chain and lifting attachment) carries the stack of records upwards and thus returns it to its initial position. The protruding contact tongues are retracted by the action of an electromagnet (incorporated in the relay box) which is energised through a micro-switch.

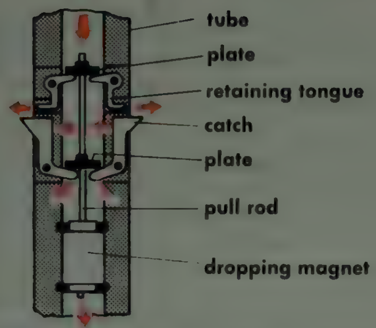


Fig. 3 SECTION THROUGH STACKER COLUMN

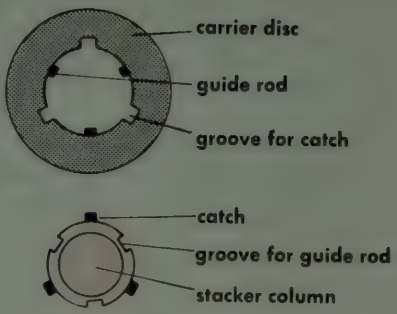


Fig. 4 CO-OPERATION OF STACKER COLUMN AND CARRIER DISC

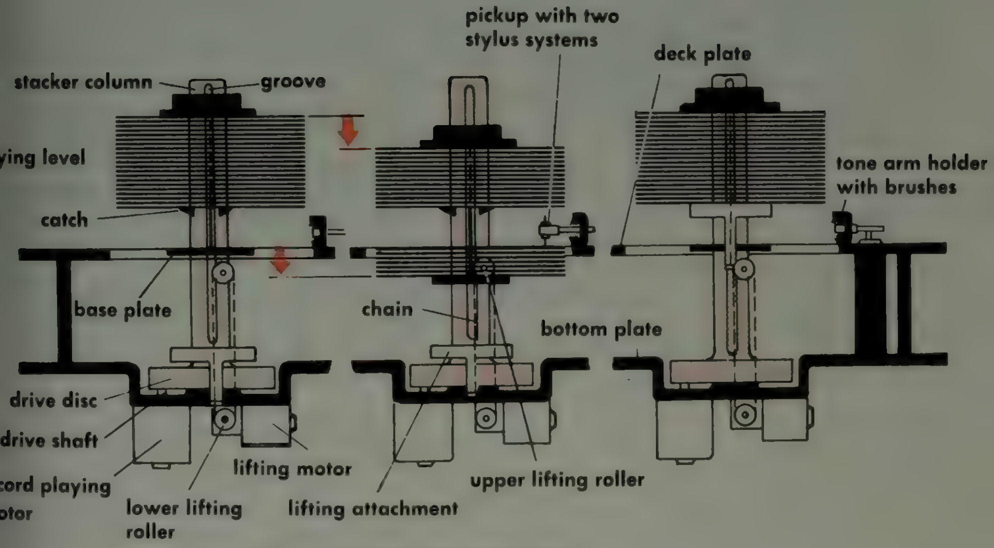


Fig. 5 RECORD PLAYING UNIT

The blast-furnace is a shaft furnace, 100 ft. or more in height, consisting of a cylindrical bottom portion (hearth with hearth bottom), from which rises an upward-widening conically tapered portion (bosh) surmounted by a taller tapered structure (shaft) which narrows towards the top where it is closed by means of a system of double conical gates (bells) (Fig. 2). This form of furnace is eminently suited to the process of ore reduction and smelting that takes place inside it. The furnace walls are lined internally with fireclay refractory bricks. The hottest part (hearth and bottom plate) has a carbon brick lining. The steel shell of the furnace is cooled. From the top, or throat, the blast-furnace is charged with metallurgical coke, iron ore and fluxes, these various materials being placed in alternate layers. Air at a pressure of around 10–20 lb./in.^{2} is blown through blast pipes into the lower part of the furnace. This air, which has first been heated to 500°–900° C in hot blast stoves (Cowper stoves), causes part of the coke to burn and form carbon monoxide. The heat of combustion raises the temperature of the coke to incandescence, drives the carbon dioxide out of the fluxes (mostly limestone) and dries the ore in the upper zones of the furnace (Fig. 1). Just above the level of the blast inlet the temperature in the furnace charge is about 1600° C. In the bosh the temperature is 1200°–1400° C; here the blast-furnace gas is composed of approximately 40% CO, 2% H_2 and 57% N_2. The rising carbon monoxide reduces some of the ore to iron and some of it to iron monoxide and escapes from the top of the furnace as "top gas" with a composition of 28–36% CO, 12% CO_2 and 52–60% N_2. The reduction process commences with the indirect reduction in the shaft at 400°–700° C, converting part of the ore to iron and part of it to iron monoxide. In the bosh, direct reduction with the glowing hot carbon in the coke at 750°–1400° C converts all the ore into metallic iron which absorbs 3–4% of carbon (in the form of elementary carbon and as iron carbide) as it trickles down into the hearth. The pig iron which collects on the hearth bottom is discharged through the tap hole at intervals of 3–6 hours and issues from the furnace at a temperature of 1250°–1450° C. It flows through the iron runner into the pig bed (sand moulds for ingots) or into the mobile ladle. Molten slag, which floats on the pig iron at the bottom of the furnace, is discharged through the slag notch and passed into moulds in which it solidifies in the form of blocks (which are used as building or paving stones) or is processed into slag wool, road stone, fertilisers, or so-called granulated slag (which is used in the manufacture of certain kinds of cement). The slag consists of lime-alumina silicates with heavy metal oxides and is formed by the silicates present in the ore and by the coke ash. Some of the top gas from the furnace is used as fuel for heating the air in the hot blast stoves or is utilised for driving large gas engines. Besides carbon, the pig iron contains manganese, phosphorus, sulphur and silicon. Grey pig iron—in which part of the carbon is precipitated as graphite—is remelted in cupola furnaces to grey cast iron, which is not particularly brittle. In white pig iron the carbon is dissolved mainly as iron carbide; this iron is brittle and hard and is processed into malleable cast iron and chilled cast iron.

A modern blast-furnace produces 500–1200 tons of pig iron per day, for a coal consumption of around 0.6 ton per ton of pig iron. The furnace also produces 200–500 tons of slag and 2000–5000 tons of top gas per day, while it consumes about 2000–5000 tons of air.

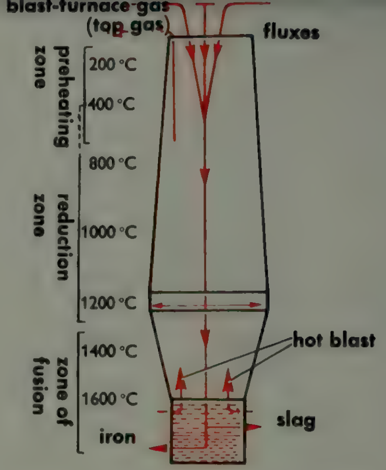


Fig. 1 ZONES IN BLAST-FURNACE

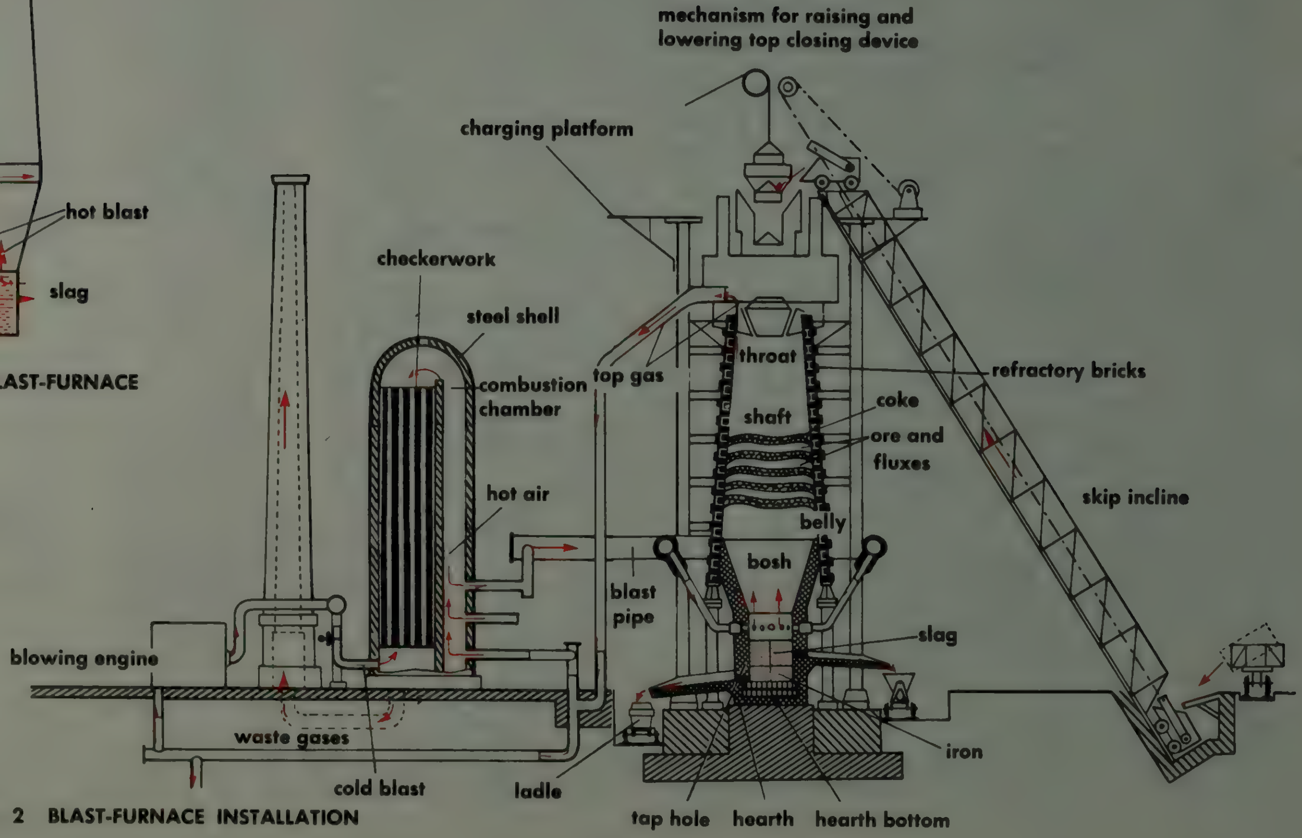


Fig. 2 BLAST-FURNACE INSTALLATION

Steel differs from pig iron in that it has a lower content of carbon. Steel-making is therefore really nothing but the removal, by combustion, of the carbon contained in the iron. As a result, the structure of the metal becomes more resilient, more flexible and—what is very important—cuttable. Steel is therefore stronger and better workable than iron.

In addition to carbon, steel contains other admixtures, e.g., sulphur and phosphorus. All these substances are removed from the pig iron by oxidation. This is achieved by bringing the molten metal into contact with air, with the result that the impurities are burned, i.e., they are transformed into their oxides by the oxygen in the air. These oxides are lighter than the molten steel and float on it as liquid slag; they must therefore be drained off before the steel itself is tapped.

Open-hearth steel-making

The central feature of a steelworks using this process is the open-hearth furnace (Fig. 1a). The pig iron, together with a certain amount of scrap iron, is deposited in the furnace hearth by a special crane. Then a gas-and-air mixture is burned over the iron, which melts at a temperature of about 1800° C. The gas burned in the furnace is preheated by a regenerative firing system. The gas, together with the air needed for combustion, enters the furnace on one side of the furnace and is preheated in the heating chamber. This mixture of gas and air is burned over the hearth; the hot waste gases flow through flues on the other side of the furnace and are discharged up the chimney. Before being discharged, however, the gases give off a considerable proportion of their heat to—initially cold—brick-lined heating chambers. The lining of these chambers consist of refractory bricks, which are heated to red-hot temperature. Then the gas flow is reversed (regenerative heating), i.e., the gas and air are admitted through the—now hot—heating chambers and absorb heat from them. This preheating of the gas and air enables the combustion temperature of the flame to be considerably raised (Fig. 1b). The air needed for the oxidation of the undesirable admixtures is provided by the combustion air. In the case of the open-hearth furnace this oxidation process is sometimes referred to as "hearth refining".

The slag produced in this steel-making process is composed of the oxides of the impurities and admixtures. It floats as a liquid on the molten steel. When it is tapped off, cooled and ground, it yields a valuable fertiliser because of its high phosphorus content. The properties of the refractory brick lining of the furnace hearth are also important, as these bricks must be able to absorb some of the sulphur and phosphorus contained in the iron.

Basic Bessemer steel-making

In this process the carbon content of the pig iron is reduced in a so-called converter—a basic Bessemer steel converter—which is an approximately pear-shaped refractory-lined steel vessel capable of holding 20–60 tons of pig iron (Fig. 2). Air is blown through the molten iron from nozzles (tuyeres) in the bottom and oxidises the carbon and other admixtures.

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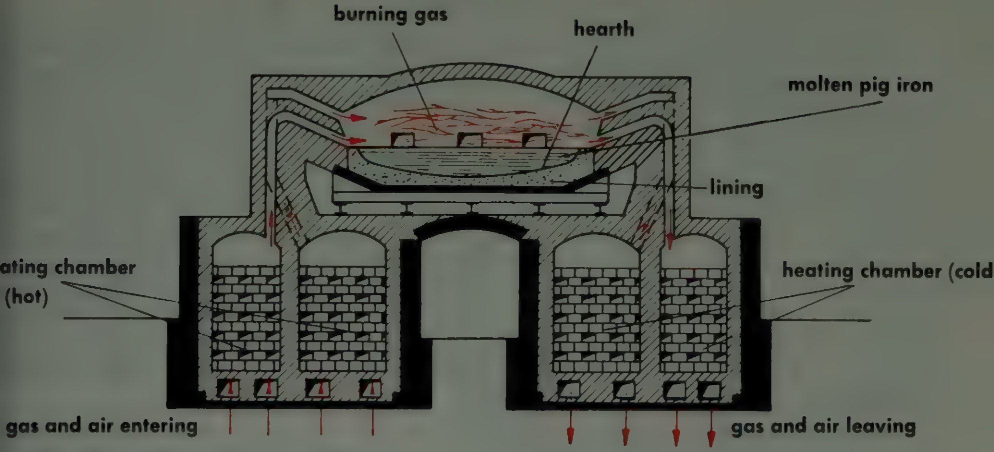

Fig. 1a OPEN-HEARTH FURNACE

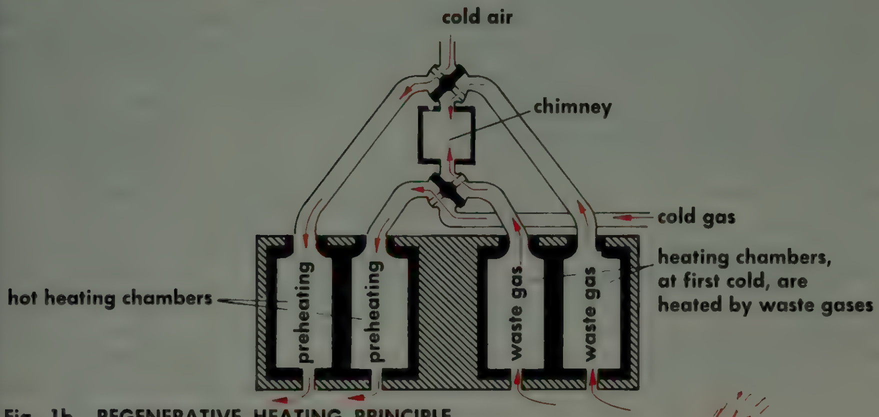

Fig. 1b REGENERATIVE HEATING PRINCIPLE

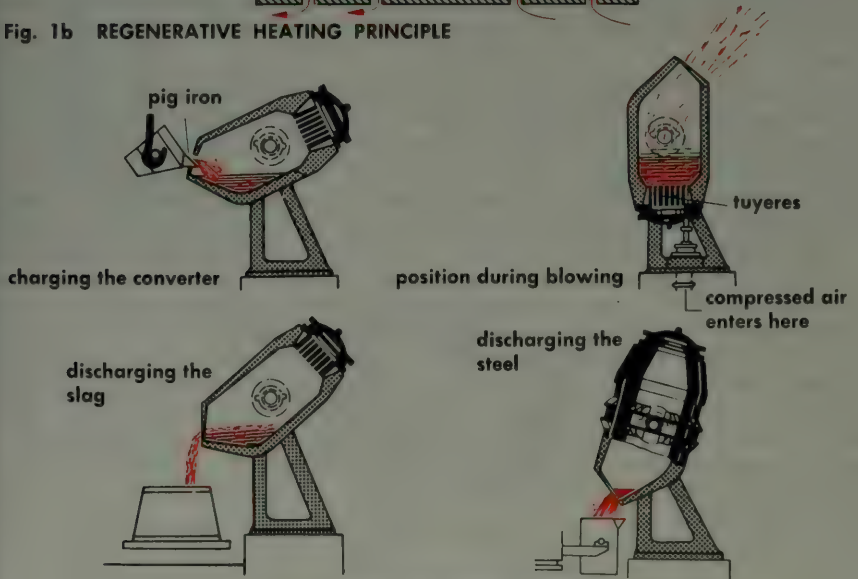

Fig. 2 BASIC BESSEMER CONVERTER

The converter is provided with trunnions and is rotatably mounted. The empty converter is tilted and filled with molten pig iron from a ladle (Fig. 2a). On completion of filling, the converter is swung to the upright position, and compressed air is blown through the pig iron from the tuyeres in the bottom (Fig. 2b).

This operation is known as blowing, and the converter steel-making process is referred to as "air refining"—as distinct from "hearth refining" in the open-hearth furnace. The oxygen in the blast air burns the carbon and the other admixtures. The lime which has been added to the converter charge promotes the separation of the oxides formed, which collect as slag on the surface of the molten metal. The latter would gradually cool and eventually solidify as a result of air-blowing, but for the fact that the combustion of the phosphorus in the pig iron produces additional heat, which not only maintains but even further raises the temperature of the molten bath. For this reason the converter method of steel-making is used chiefly for pig iron with a high content of phosphorus.

When oxidation has been completed (this is judged from the colour of the flames emerging from the top of the converter), the converter is tilted, the molten slag is first poured off, and then the molten steel—mild steel—is discharged into ingot moulds (Figs. 2c and 2d).

The conventional Bessemer—as distinct from the "basic" Bessemer—steel-making process operates on the same principle, except that the converters employed are usually smaller and the refractory lining is of a different composition, the pig iron treated in this process having a low content of phosphorus and sulphur.

Manufacture of electric furnace steel

Another steel-making process, which yields mild steel of exceptionally high purity, is the electric process. The requisite heat is in this case not produced by the burning of gas or coal, but by an electric current. The raw material for the process is usually molten open-hearthed steel, which is thus further purified and refined in quality. However, the electric furnace can alternatively be charged with cold scrap iron with a certain quantity of added pig iron. In most electric furnaces the requisite heat is produced by an electric arc whicn is formed between a number of carbon electrodes and the surface of the molten bath (p. 118, vol. I). No combustion air for oxidising the undesirable admixtures is supplied to the electric arc furnace (Fig. 3); instead, iron oxides are added, which give off their oxygen.

Oxygen steel-making

The application of pure oxygen in the refining process has, in recent years, led to the emergence of new steel-making methods. The steels made by these methods are as good as open-hearth steels. A feature common to all of them is that almost pure oxygen—produced in large quantities by so-called "tonnage oxygen" plants—is blown onto or into the molten pig iron. In the LD process (Fig. 4a) a jet of oxygen is blown at high pressure through a tube (called a lance) onto the surface of molten iron in a converter-type vessel. An offshoot, differing chiefly in the technique employed, is the LD–AC process, which can deal with iron having a higher phosphorus content, the converter being additionally charged with lime. In the OLP process powdered lime is injected with the stream of oxygen. The rotor process (Fig. 4b) uses a long cylindrical kiln which is slowly rotated (about $\frac{1}{2}$ r.p.m.). A primary oxygen jet blows oxygen into the molten bath itself, and a secondary oxygen nozzle above the bath ensures heat economy by burning the Co generated to CO_2. In the Kaldo process (Fig. 4c) a converter rotating on an inclined axis at speeds up to 30 r.p.m. is used; an oxygen jet is blown against the surface of the bath.

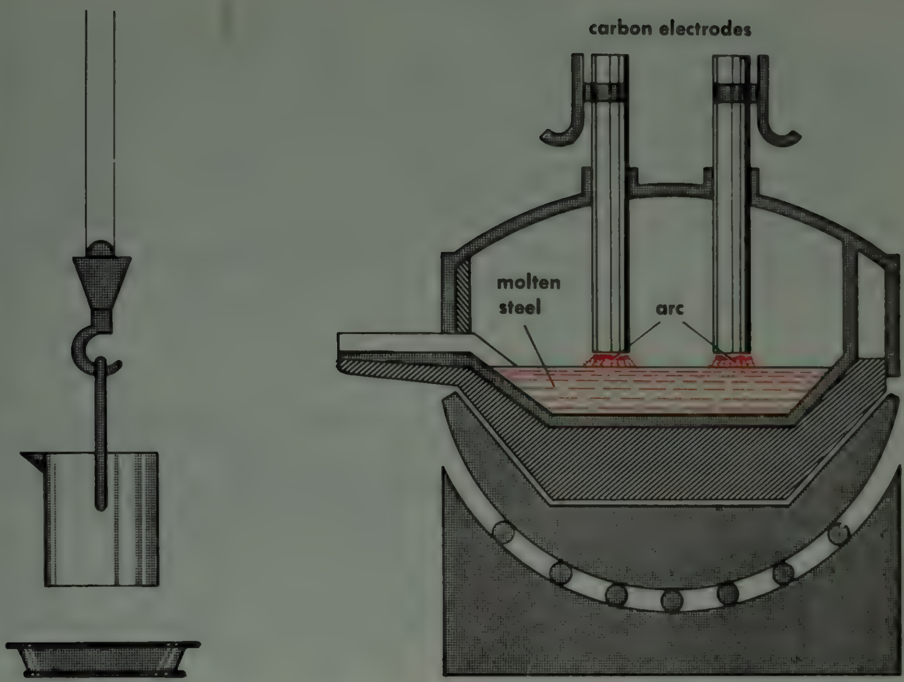

Fig. 3 ELECTRIC ARC FURNACE

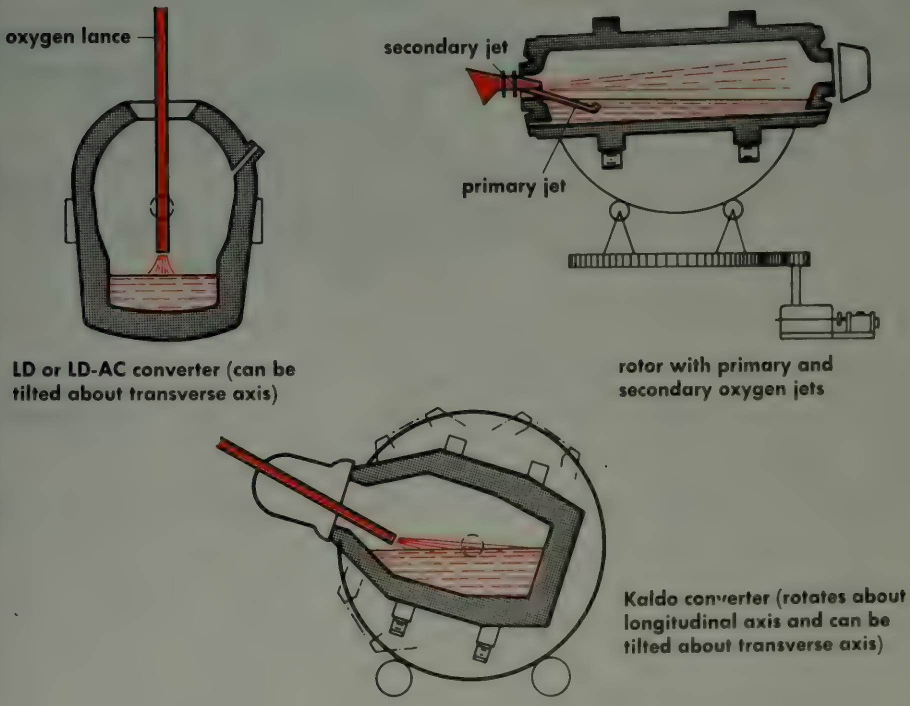

Fig. 4 VARIOUS OXYGEN STEELMAKING PROCESSES

In a rolling mill, ingots (pieces of steel of square cross-section) and slabs (of rectangular cross-section) are rolled, i.e., they are passed between rolls whereby they undergo an increase in length and a reduction in height (or depth). The white-hot ingot of steel is inserted between two rotating rolls made of cast iron or cast steel. The gap between these rolls is smaller than the thickness of the ingot. The top and bottom rolls are separately driven. Quite often they are both driven by one and the same electric motor, but in that case the drive is divided in a reduction gear unit interposed between the motor and the rolls. The thickness of the rolled ingot is varied by adjustment of the top roll (Fig. 1). The ingot or slab which is fed to the rolls is gripped by them and pulled through the gap between the top and the bottom roll by frictional action. The cross-section of the ingot is thereby reduced, and at the same time it undergoes an increase in length. The rolls used in rolling mills present a wide variety of shapes, depending on the type of rolled section to be produced. The rolls of sheet mills are smooth. The slabs (in smaller sizes they are known as "sheet bars") are rolled so as to increase their width. Of course, it is not possible to reduce a large slab of steel to a $\frac{1}{16}$ in. sheet in a single rolling operation. The reduction in thickness has to be effected in several successive operations, the slab either being passed again and again between the same pair of rolls, whose distance from each other is progressively reduced, or between a series of pairs (with gaps of diminishing width) installed one behind the other. An installation of this kind is called a rolling train. For producing rolled steel structural sections and the like, the ingots are first rough-rolled and then passed between a series of grooved rolls (Figs. 2a–2c). The shape of the grooves successively approximates to the final shape of the rolled section. I-sections, angles, bulb sections, etc. are produced in this way (Fig. 3).

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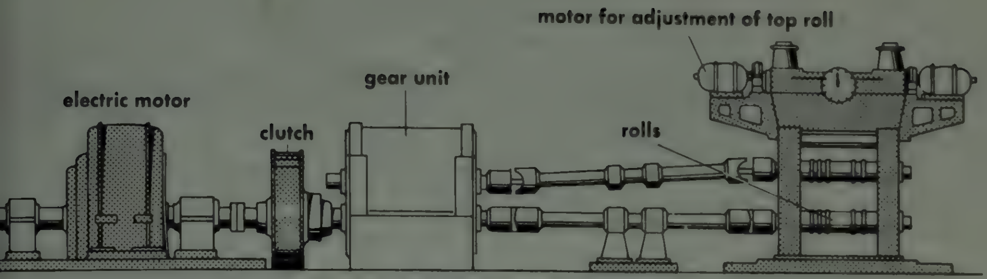


Fig. 1 ROLLING MILL WITH DRIVE MACHINERY

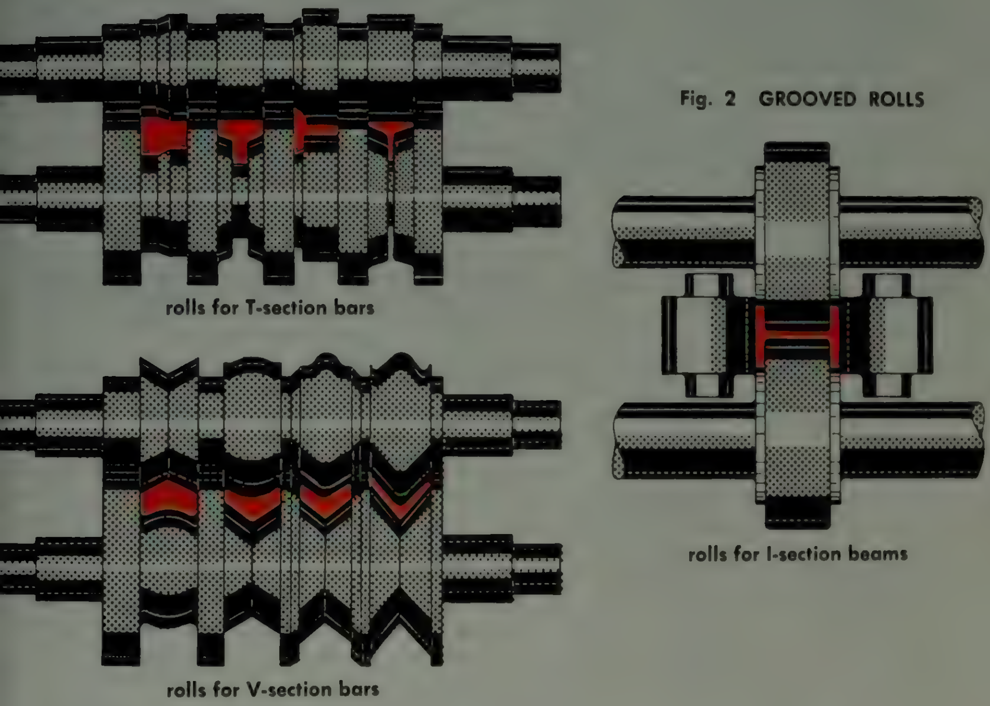


Fig. 2 GROOVED ROLLS

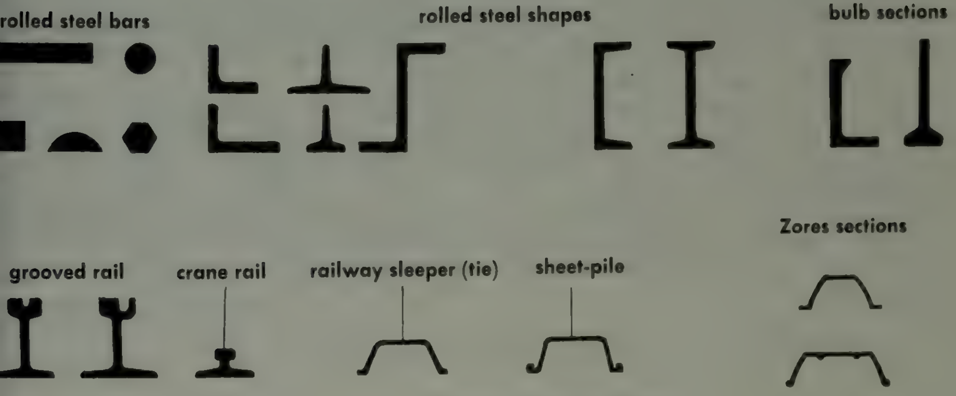


Fig. 3 ROLLED STEEL SECTIONS

Broadly speaking, steel tubular products (pipes) may be subdivided into: welded pipes and seamless pipes. Welded pipes are produced from steel strip which is bent to a tubular shape and whose edges are then joined by welding. Seamless pipes are produced from cast or rolled round billets at rolling temperature. Depending on the pipe diameter and wall thickness required, a number of different methods of manufacture are used for both types of pipe.

The oldest continuous pipe-welding process is the Fretz-Moon process, based on the principle of "forge welding". An endless strip of steel (called the skelp) is passed through a furnace in which it is heated to about 1400° C and is then fed through shaping rolls which bend the skelp to a tubular shape (Fig. 1). It next passes through welding rolls which press the edges of the longitudinal slot together, so that they become welded to each other by their own heat. In other processes for welded tube manufacture the skelp is shaped in the cold condition, only the edges of the slot being heated to enable the weld to be formed. Various methods of electric arc welding are used, one of the most common being the submerged-arc method (in which the welding is shielded by a blanket of granular fusible material, or flux). Induction welding and resistance welding processes are also widely employed, in which electric induction and electric resistance are respectively utilised for heating the edges. In resistance welding the electrodes are copper discs, on each side of the opening to be welded. A special kind of pipe is the spiral-wound pipe made of steel strip wound and welded with a continuous spiral seam.

The oldest method of manufacturing seamless tubular products is by means of the Mannesmann piercing process, which employs the principle of helical rolling. The machine comprises two steel rolls whose axes are inclined in relation to each other. They both rotate in the same direction. The space between the rolls converges to a minimum width called the gorge. Just beyond the gorge is a piercing mandrel. A solid round bar of steel is heated to about 1300° C and, revolving in the opposite direction to the rolls, is introduced between the rolls (on the left in Fig. 2a). When the leading end of the bar has advanced to the gorge, it encounters the mandrel, which thus forms the cavity in the bar as the latter continues to move through the rolls (Figs. 2b, c). The thick-walled tube produced by this process can subsequently be reduced to thin-walled tube by passing it through special rolls in a so-called Pilger mill. These rolls vary in cross-sectional shape round their circumference. The tube, fixed to a mandrel, is gripped by the narrow part of the rolls (Fig. 3a) and its wall thickness is thereby reduced (Figs. 3a–c) until the rolls have rotated to such an extent that the wider part of their cross-section is reached and the tube is thus no longer gripped. The tube is then pulled back some distance, so that again a thick-walled portion is gripped by the rolls. The mandrel is rotated at the same time in order to ensure uniform application of the roll pressure all round the tube.

In the Stiefel piercing process for seamless tube manufacture a round bar is first pierced on a rotary piercing mill and the heavy-walled shell obtained in this way is then reduced in a second piercing operation, on a two-high rolling stand, to a tube with a thinner wall (Fig. 4).

In the rotary forge process a square ingot, heated to rolling temperature, is shaped to a shell closed at one end. This shell is then reduced and stretched on a rotary piercing mill and finally passed through sets of four rolls whereby the diameter is progressively reduced (Fig. 5).

The latest method of seamless tube manufacture is by hot extrusion: a hot billet is forced through a die provided with a mandrel for piercing the hole. In this way the tube of the correct final diameter and wall thickness is produced in a single operation.

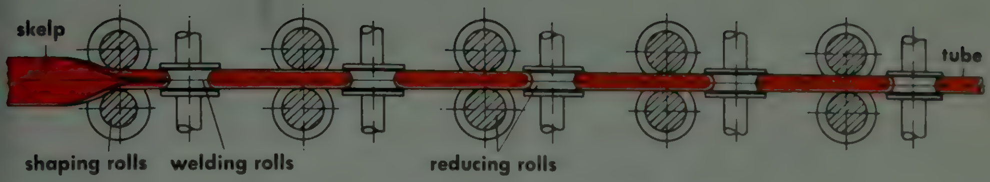


Fig. 1 FRETZ-MOON PROCESS FOR WELDED TUBE MANUFACTURE

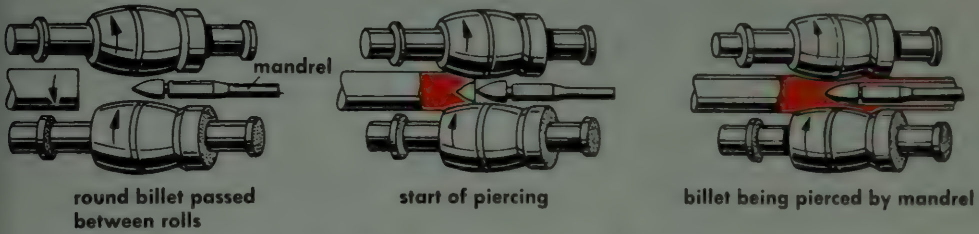


Fig. 2 ROTARY PIERCING MILL FOR SEAMLESS TUBE MANUFACTURE (MANNESMANN PROCESS)

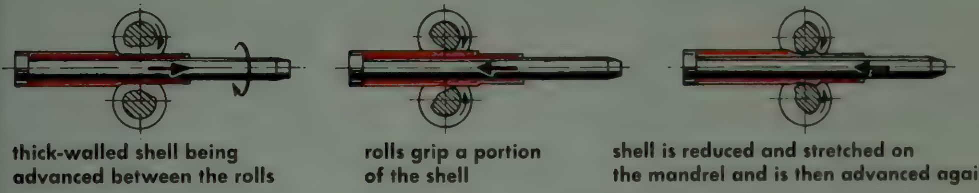


Fig. 3 PILGER STEP-BY-STEP ROLLING PROCESS

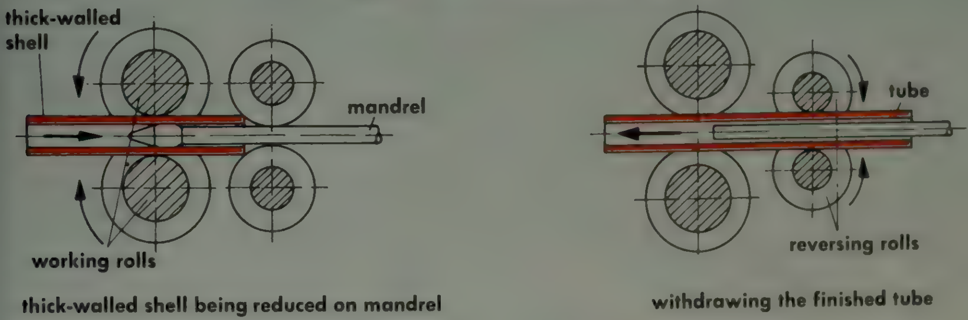


Fig. 4 STIEFEL PIERCING PROCESS

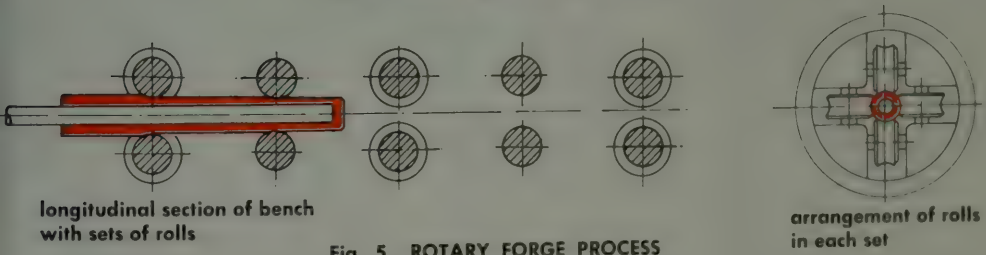


Fig. 5 ROTARY FORGE PROCESS

Manufacture of steel tubes. Sulzer, Winterthur, Switzerland
Photo Roland Schneider, Len Sirman Press

The preliminary treatment of the material to be manufactured into wire is done in the rolling mill (see page 38). Billets (square section ingots) about 7 cm ($2\frac{3}{4}$ in.) thick are rolled to round wire rod of about 5 mm (0.4 in.) diameter. The action of atmospheric oxygen causes a coating of mill scale to form on the hot surface of the rod and must be removed. This descaling can be done by various mechanical methods (e.g., shot-blasting) or by pickling, i.e., immersion of the wire rod in a bath of dilute sulphuric or hydrochloric acid. After pickling, the wire rod may additionally undergo a jolting treatment on the device illustrated in Fig. 1, which operates on the drop-hammer principle and dislodges the scale loosened by the acid. The remaining acid is removed by immersion of the wire rod in lime water.

The actual process of forming the wire is called drawing and is carried out on the metal in the cold state—not hot, as in rolling. The wire is pulled through a draw plate (Fig. 3a), which is a steel plate provided with a number of holes (dies) of various diameters. These dies have holes which taper from the diameter of the wire that enters the die to the smaller diameter of the wire that emerges from the die. Each passage through a die reduces the diameter of the wire by a certain amount. By successively passing the wire through dies of smaller and smaller diameter, thinner and thinner wire is obtained (Fig. 3b). To pass a wire through a die, the end of the wire is sharpened to a point and threaded through the die. It is seized by a gripping device and rapidly pulled through the die. This is assisted by lubrication of the wire. The actual minimum diameter (d_2) of the die is smaller than the final diameter (d_3) of the wire, as the elasticity of the metal causes it to expand to d_3 (Fig. 3). Copper and brass wire is manufactured from strips of sheet metal whose sharp edges are rounded by means of rollers. These strips are then drawn through dies (Fig. 2). The thick wire rod is coiled on a vertical spool called a swift and is pulled through the die by a rotating drum mounted on a vertical shaft which is driven by bevel gearing. The drum can be disconnected from the drive by means of a clutch.

The dies used in the modern wire industry are precision-made tools, manufactured in tungsten carbide for larger sizes or diamond for smaller sizes. Tungsten carbide dies are more accurate and much longer lasting than steel draw plates or steel dies.

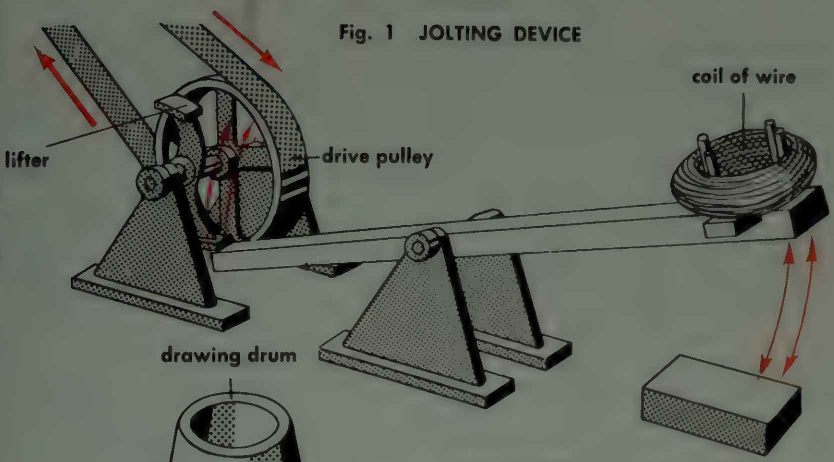


Fig. 1 JOLTING DEVICE

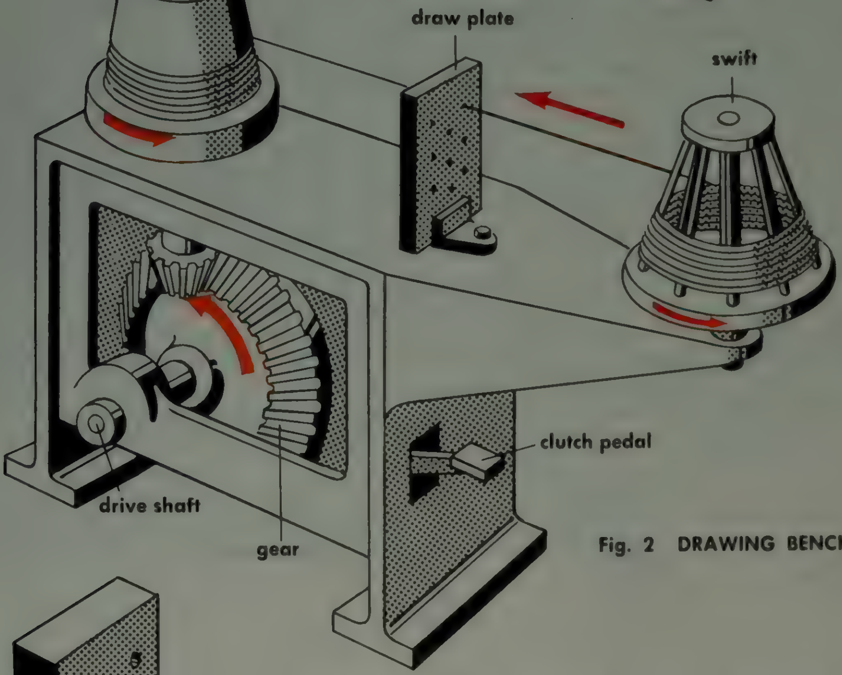


Fig. 2 DRAWING BENCH

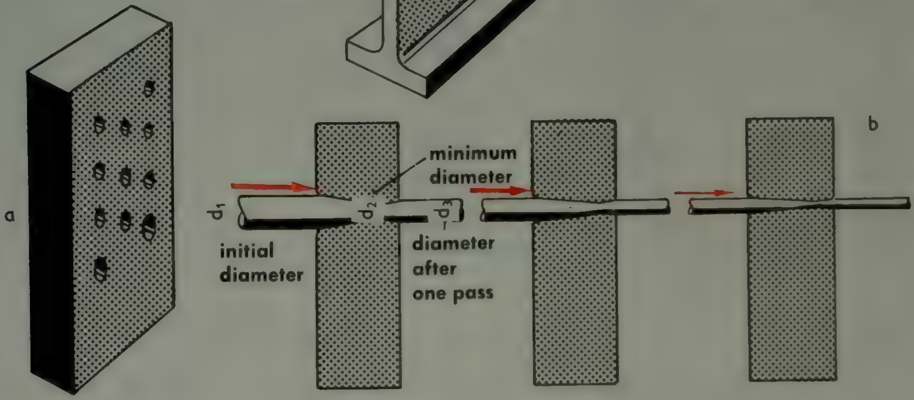
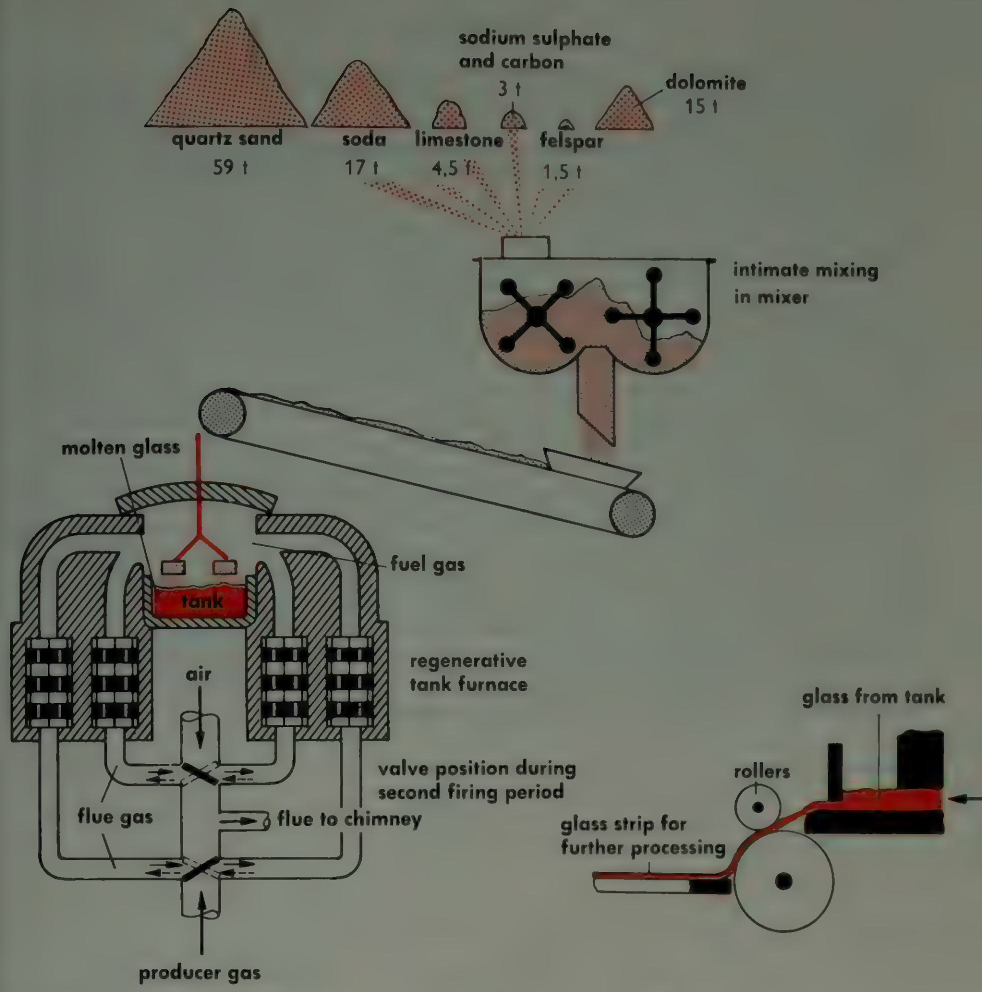


Fig. 3 PROGRESSIVE REDUCTION OF WIRE DIAMETER IN DIES

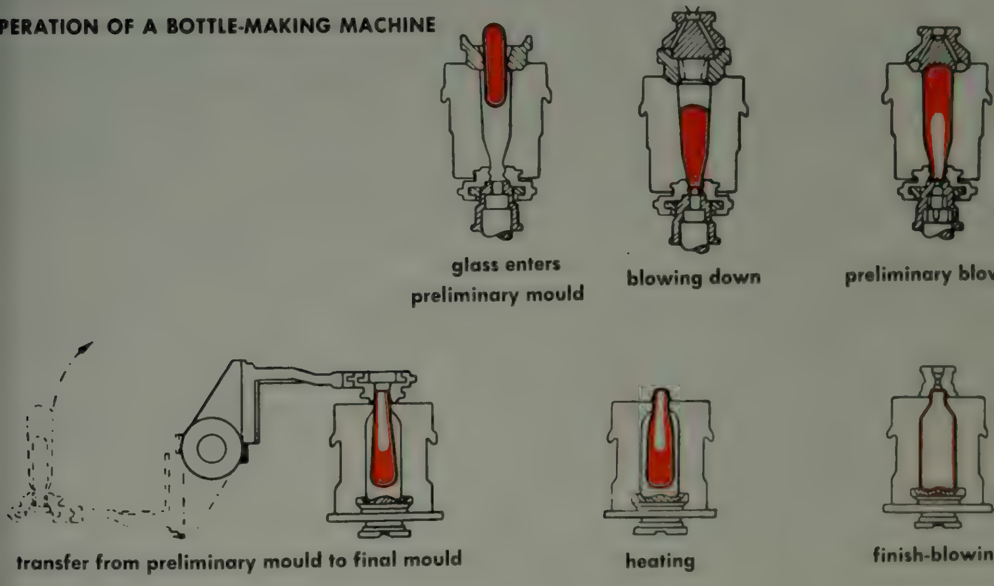
The raw materials for glass manufacture are now still—as they were in ancient times—sand, soda and limestone or chalk. From the chemical point of view glass consists of compounds of silicates with alkaline and alkaline-earth oxides. After melting, glass with the correct composition and subjected to the right treatment (cooling), must not become crystalline on solidifying.

There are hundreds of recipes for making glass. One of these is the following: 59 parts of quartz sand, 17 parts of soda, 15 parts of dolomite, $4\frac{1}{2}$ parts of limestone, 3 parts of sodium sulphate and carbon, and $1\frac{1}{2}$ parts of felspar. These ingredients are intimately mixed together. 20 to 30 per cent of cullet (broken waste glass) is added to this mixture, which is then melted in tank furnaces with capacities of up to 1500 tons of material, or in pot furnaces containing a number of pots, each holding up to about 2 tons of material. First, the fluxes, e.g., soda and the cullet added to the mixture, are melted. The melting soda forms low-melting alkali silicates with the sand; these silicates enter into further reaction with the high-melting constituents and form the final glass melt. The carbon dioxide of the carbonates is expelled during melting and also at the end of the melting process, when the temperature is raised somewhat. Thereafter the temperature is reduced from the 1400°–1500° C, at which melting is carried out, to a temperature of 900°–1200° C. Undesirable discoloration of the glass is removed in various ways, e.g., by addition of manganese dioxide, antimony oxide and arsenic. Coloured glass is produced by the addition of metallic oxides (iron, copper, manganese, chromium, etc.). Glass melting furnaces are heated with gas or oil. The gas and the combustion air are passed through a heat exchanger; this raises their temperature before combustion takes place in the furnace (regenerative furnaces). Depending on the composition, the thickly liquid melted glass is formed to the required shapes by casting, rolling, drawing, moulding, blowing, spinning or pressing, either with or without blowing. These processes are generally applied by machines. Hand forming processes, usually in conjunction with blowing with blowpipes, are now relatively little used. After forming has been done, it is most essential that cooling be correctly applied, so as to keep the glass as free as possible from stresses. The sheet glass, rolled glass, pressed glass, hollow glassware, etc. can be further worked and manufactured by cutting, drilling, milling, cementing, welding, bending, etching, grinding, engraving, polishing, enamelling, etc.

Special glasses for technical and scientific purposes (e.g., filter glasses, refractory glasses, chemical-resisting glasses, malleable glasses, etc.) are produced by the use of additives to the standard recipe or are made from quite different raw materials from those mentioned above. These special kinds of glass are usually manufactured in fairly small pots (which are constructed of aluminium oxide or some other highly refractory material).



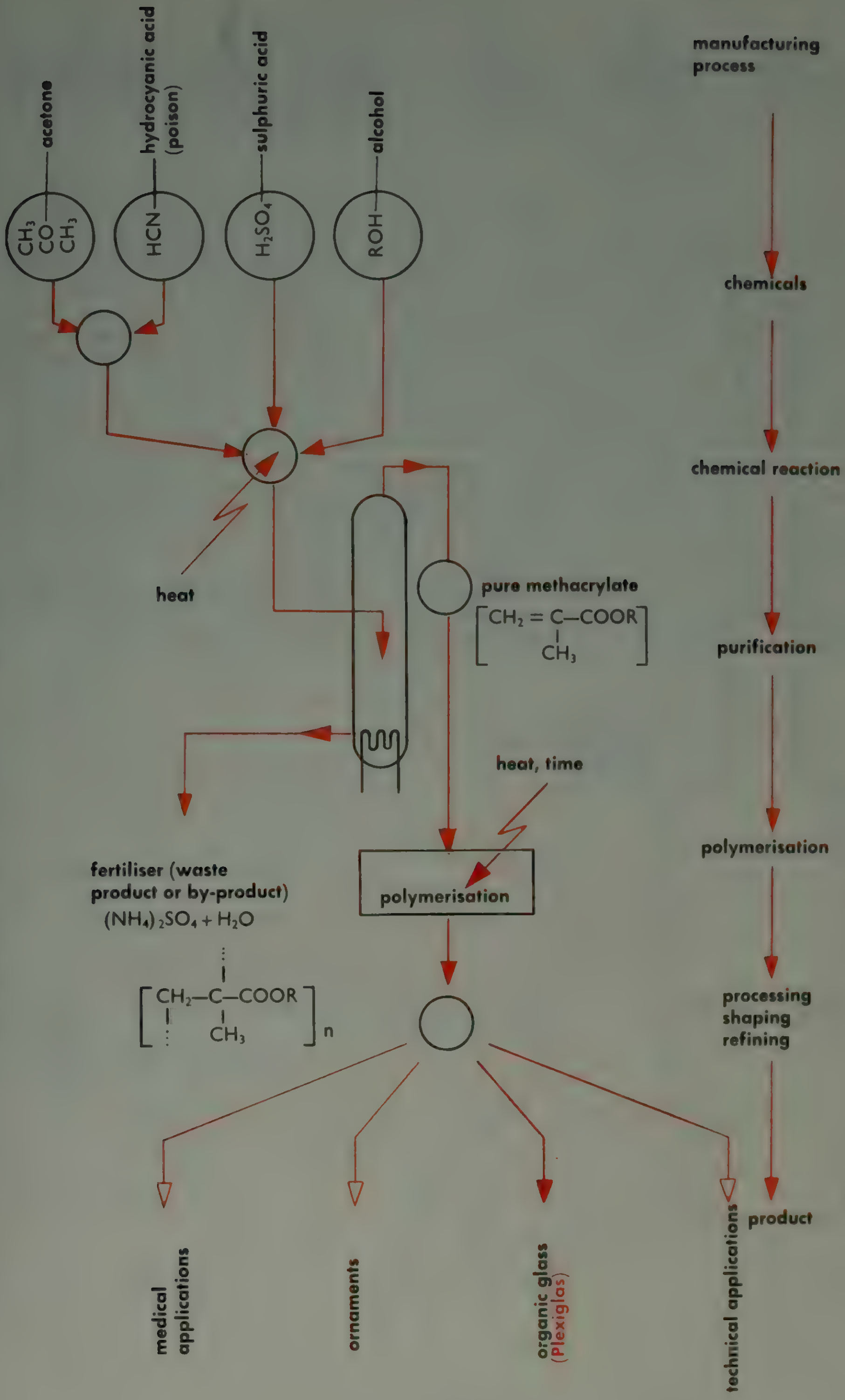
PERATION OF A BOTTLE-MAKING MACHINE



Plexiglas is a synthetic "organic glass" which is produced in large quantities by the chemical industry and which has two major advantages over inorganic glass (see page 46): it weighs only half as much and it is very easy to work. Plexiglas is manufactured from acetone, hydrocyanic acid, sulphuric acid and alcohol and may be polymerised in various ways, whereby it becomes hard and insoluble. Plexiglas has even better transparency than ordinary window glass; ultraviolet rays and X-rays pass through it unhindered, but heat rays are stopped by it. It can be produced in any desired shape (solution, powder, plates, blocks, tubes, etc).

Plexiglas can be drilled, turned, filed, cast, sawn, ground, punched, welded, cut and polished. A further important property is its completely non-toxic character (as a finished product), although it is made from hydrocyanic acid (prussic acid), which is a highly poisonous substance. It is finding increasingly widespread application in surgery: artificial eyes, ears, noses, fingers, entire hands and limbs, dentures, etc. are made from Plexiglas. The human tissues coalesce with the synthetic parts, so that it is even possible to replace defective heart valves by ones made of Plexiglas. Because of its good stability to light and transparency, Plexiglas is used for many optical purposes. Many physical instruments, measuring instruments and models which have to be transparent are made from Plexiglas (protective goggles, fluorescent tubes, etc.). This material is also extensively used in the manufacture of musical instruments, ornaments and toys. Also, it is an important material in aircraft and motor engineering and is furthermore used in building construction. Plexiglas is resistant to temperatures of 100° C (and even higher); it is also resistant to water, caustic alkalies, dilute acids, petrol,[1] and mineral oils. However, it swells in alcohols, esters, ketones, benzene (and other aromatic hydrocarbons), and chlorinated hydrocarbons. Plexiglas can burn at high temperatures.

1. Gasoline in U.S.A.



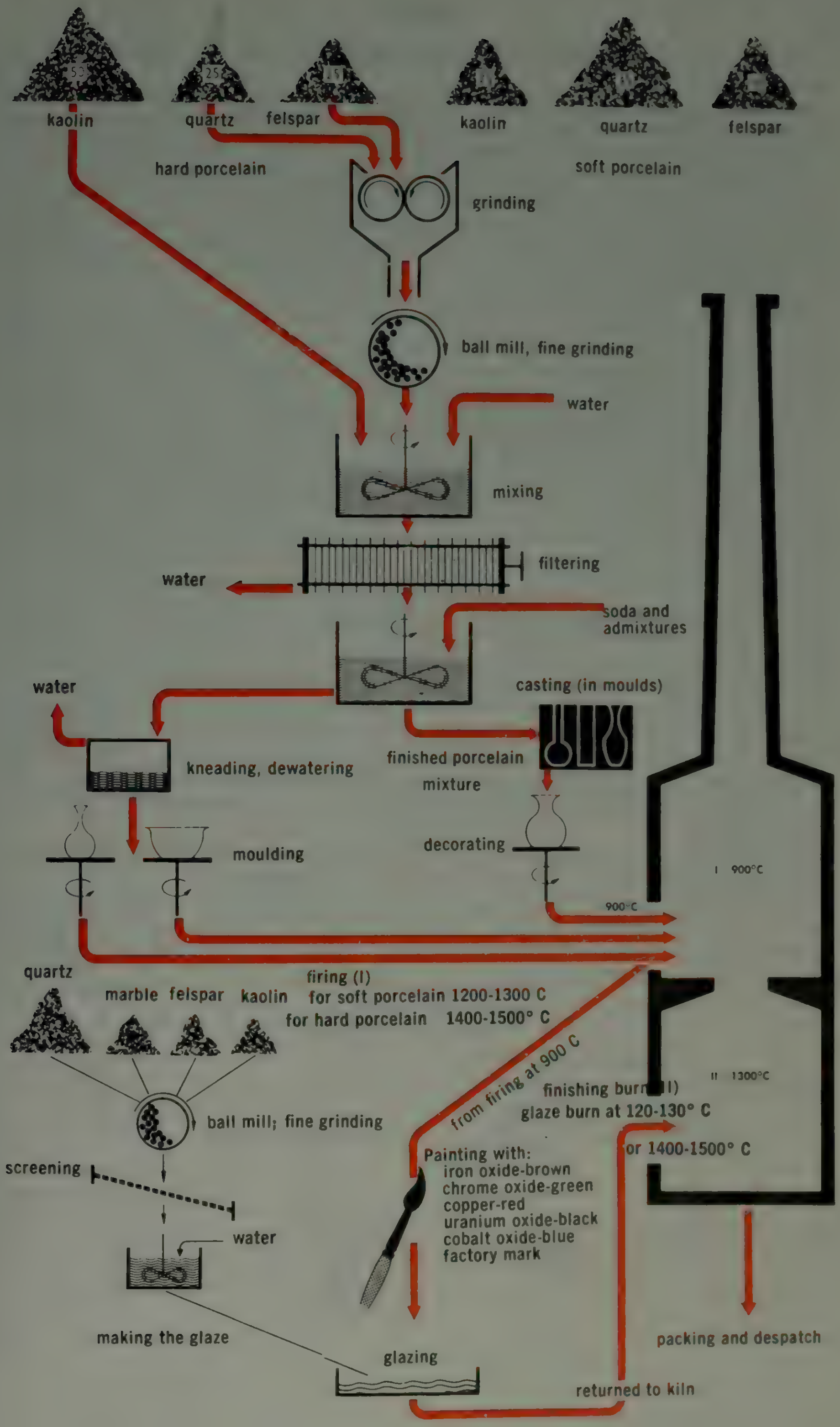
manufacturing process
chemicals
chemical reaction
purification
polymerisation
processing shaping refining
product
acetone
hydrocyanic acid (poison)
sulphuric acid
alcohol
CH_3 CO CH_3
HCN
H_2SO_4
ROH
heat
pure methacrylate
$\left[CH_2 = \underset{\displaystyle CH_3}{\underset{|}{C}}-COOR\right]$
heat, time
polymerisation
fertiliser (waste product or by-product)
$(NH_4)_2SO_4 + H_2O$
$\left[CH_2-\underset{\displaystyle CH_3}{\underset{|}{C}}-COOR\right]_n$
medical applications
ornaments
organic glass (Plexiglas)
technical applications

Porcelain is a translucent non-porous ceramic material. It is made by firing a mixture of kaolin (porcelain earth), quartz and felspar. Hard porcelain is usually made from a mixture comprising 50% kaolinite, 25% quartz and 25% felspar; for soft porcelain the corresponding proportions are 25%, 45% and 30%. Soft porcelain is fired at lower temperatures and is therefore cheaper to manufacture.

Manufacturing procedure:

Kaolinite is a clay mineral which is present in kaolin. The latter material is washed, screened and thoroughly mixed with finely ground quartz and felspar, in conjunction with the addition of water. Soda and other additives are added. The plastic porcelain mixture is kneaded and moulded or cast in plaster moulds. After it has dried, it is fired in a kiln at about 900° C. The porcelain has then become hard and watertight, but dull. This treatment is followed by glazing. The objects are dipped into a liquid containing the same ingredients as the material, but in which quartz and felspar predominate. They are then fired at a temperature of about 1400° C. The glazing thereby becomes a continuous, firmly adherent coating. In this heating treatment, which lasts from 20 to 30 hours, the fine raw material particles of the porcelain mixture are sintered (partially fused) together. Coloured decorative patterns can be produced by painting the glazed porcelain with dyes. In order to make these melt together with the glazing and become durable, the painted objects are heated in an enamelling furnace to about 800° C. In modern ceramic technique the paint is, alternatively, sometimes applied before the glazing.

Porcelain is used for household purposes, in laboratories, in industry (more particularly as an insulating material in electrical engineering). It is impermeable to liquids and gases and is very largely unaffected by temperature variations.



kaolin
quartz
felspar
kaolin
quartz
felspar
hard porcelain
soft porcelain
grinding
ball mill, fine grinding
water
mixing
filtering
water
soda and
admixtures
water
casting (in moulds)
kneading, dewatering
finished porcelain
mixture
moulding
decorating
900°C
900°C
1300°C
quartz
firing (I)
marble felspar kaolin
for soft porcelain 1200-1300 C
for hard porcelain 1400-1500° C
from firing at 900 C
finishing bur
glaze burn at 120-130° C
or 1400-1500° C
ball mill; fine grinding
screening
Painting with:
iron oxide-brown
chrome oxide-green
copper-red
uranium oxide-black
cobalt oxide-blue
factory mark
water
making the glaze
glazing
packing and despatch
returned to kiln

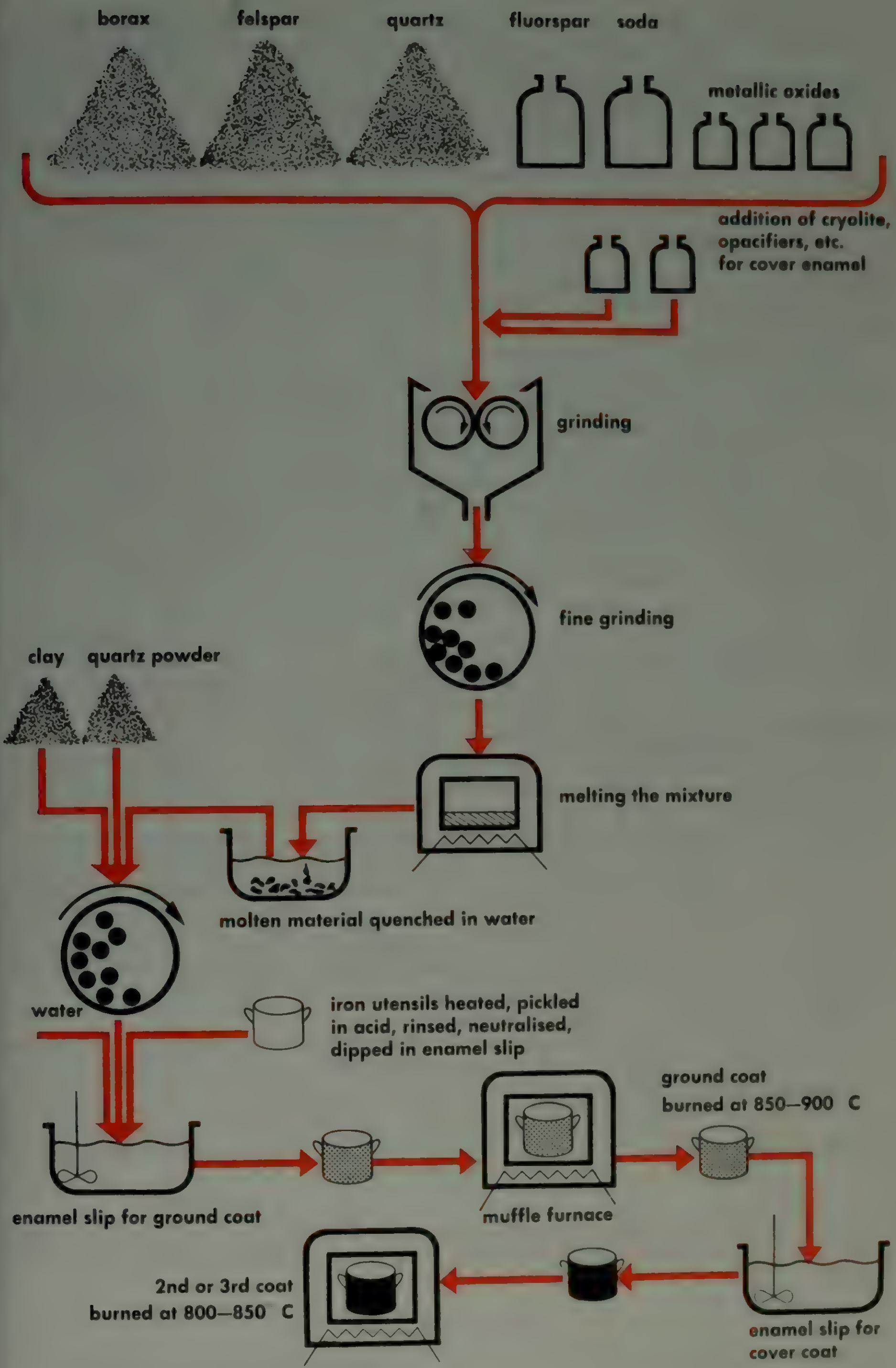
Enamel is a vitreous glaze, of inorganic composition (chiefly oxides), fused on a metallic surface.

Glass (see page 46) is particularly resistant to corrosion by atmospheric influences and chemicals and has a smooth and very strong surface. But glass is fragile. When the good properties of glass are combined with the strength of steel or cast iron, the objects made from these materials (kitchen utensils, baths, pipes, basins, etc.) have excellent service properties. The name "vitreous enamelling" or "porcelain enamelling" is applied to such materials. In particular cases the two materials supplement each other so well that entirely new material properties are obtained: jet engines, the internal surfaces of certain parts of marine propulsion engines, etc. are enamelled in order to make the surfaces resistant to high temperatures.

Enamel has been known since ancient times, when it was used (as it still is) for ornamental purposes on precious and non-ferrous metals. In the last few hundred years, however, it has been used chiefly for improving the surface properties of steel and cast iron objects and protecting them against corrosion. An enamel consists of glass-forming oxides and oxides that produce adhesion or give the enamel its colour. A normal enamel may consist, for example, of 34 (23) parts of borax, 28 (52) parts of felspar, 5 (5) parts of fluorspar, 20 (5) parts of quartz, 6 (5) parts of soda, 5 (2.5) parts of sodium nitrate, 0.5–1.5 parts of cobalt, manganese and nickel oxide, (6.5) parts of cryolite. The figures not in parentheses relate to a ground-coat enamel, while those in parentheses relate to a cover enamel, to which 6–10% of an opacifier (a substance which makes the enamel coating opaque, e.g., tin oxide, titanium silicate, antimony trioxide) and a colour oxide is added. This mixture is ground to a very fine powder and melted; the hot melt is quenched by pouring it into water, and the glass-like "frit" that is thus produced is ground fine again. During grinding, water (35–40%), clay, and quartz powder are added; opacifiers and pigments may also be added. The enamel "slip" (thick slurry) obtained in this way must be left to stand for a few days before use. The metal objects to be enamelled are heated thoroughly, pickled in acid, neutralised in an alkaline bath, and rinsed. Next, the ground-coat enamel slip is applied to them by dipping or spraying and is fired at 850°–900° C, so that it fuses to form a glass coating. The ground-coated objects are then provided with one or more coats of cover enamel, each coat being fired at 800°–850° C in a muffle furnace.

As an enamel coat is always more brittle than the underlying metal, the enamel will crack or spall if the object is deformed or roughly knocked.

From the chemical point of view enamel is a melted mixture of silicates, borates and fluorides of the metals sodium, potassium, lead and aluminium. Colour effects are produced, for example, by the admixture of various oxides to the melt (oxides of iron, chromium, cadmium, cobalt, nickel, gold, uranium and antimony).


borax
felspar
quartz
fluorspar
soda
metallic oxides
addition of cryolite,
opacifiers, etc.
for cover enamel
grinding
fine grinding
clay
quartz powder
melting the mixture
molten material quenched in water
water
iron utensils heated, pickled
in acid, rinsed, neutralised,
dipped in enamel slip
ground coat
burned at 850—900 C
enamel slip for ground coat
muffle furnace
2nd or 3rd coat
burned at 800—850 C
enamel slip for
cover coat

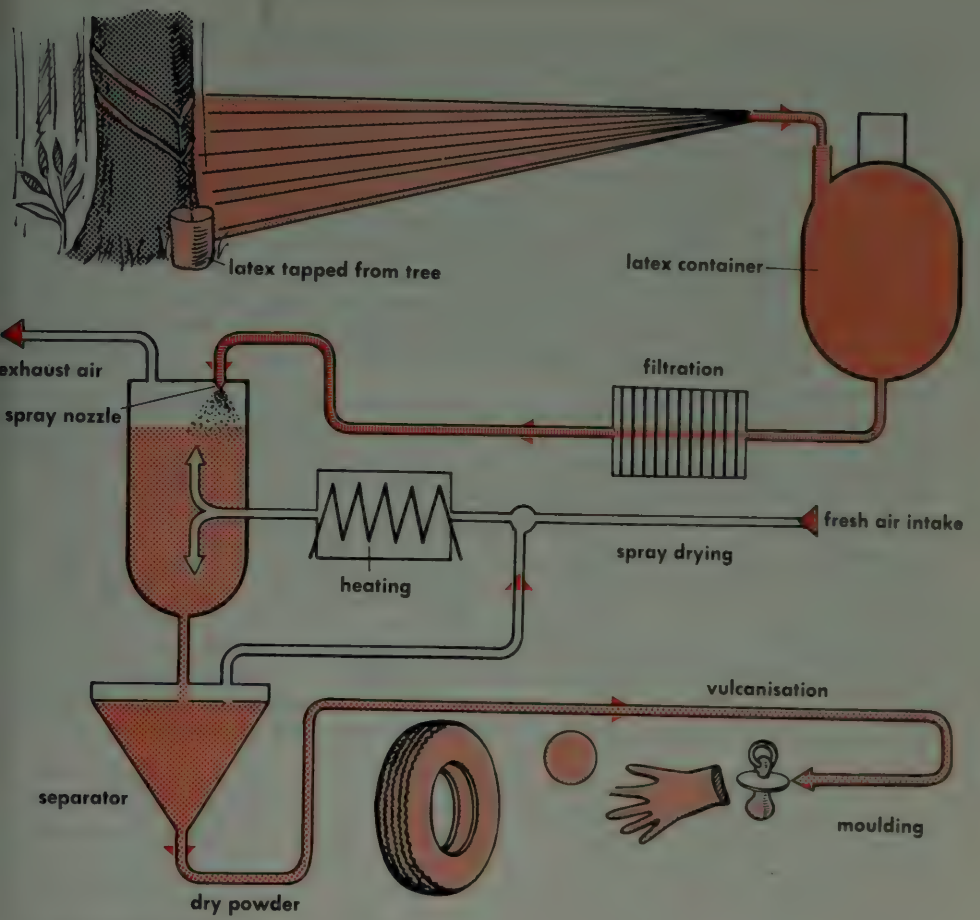
Rubber or indiarubber is the name applied to the vulcanised (see page 58) final product of a natural vegetable gum (caoutchouc). Natural rubber is present in the form of tiny droplets in the juice (latex) of the rubber tree (Hevea brasiliensis) which attains a height of 60–80 ft. and is grown in plantations in tropical countries. Tapping the trees for latex consists in removing a shaving of bark with a sharp knife. The cut passes through the latex tubes and there is a flow of latex in consequence. Most trees yield about 5 lb. of rubber per annum on an average. To transform the milky latex into a serviceable product, it is necessary to remove the water by spraying and drying, or by acidification, coagulation, washing and rolling, or the latex may be coagulated by treating it with smoke. The two most important forms of plantation rubber are sheet and crepe. Sheet is generally dried in smoke and is dark brown in colour. Crepe, which is air-dried, has a much lighter colour and is passed through heavy rollers prior to drying.

The raw rubber is cleaned, chemicals and fillers are added to it, and it is finally vulcanised with sulphur. As a result, a highly elastic product is obtained. This property is due to the structure and form of the rubber molecule. Pure rubber (caoutchouc) is a compound containing carbon and hydrogen only. It consists of long chains of interconnected isoprene molecules. A rubber molecule contains upwards of two thousand of such elementary units joined together in a linear arrangement. The vulcanised polyisoprene molecule has an angular pattern and polar groups. When a filament of rubber is stretched, the positions and angles of the entangled individual rubber molecules are temporarily changed. When the rubber is released, the molecules spring back to their original shape. These changes in shape involve no permanent mechanical changes and can be repeated indefinitely or, at any rate, until the rubber fractures for other reasons (e.g., aging due to the action of oxygen and exposure to strong light). Rubber does not acquire its high degree of elasticity until it has been vulcanised, whereby the rubber molecules become slightly interlinked by "sulphur bridges". Untreated raw rubber is very sensitive to heat and would, in that condition, be unsuitable for the manufacture of, for example, motor tyres.[1]

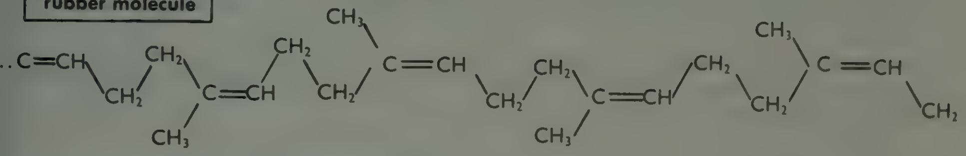
Because of its extraordinary elastic properties, rubber is a raw material for a vast range of products (something like fifty thousand different applications). It is only since about the middle of the last century that rubber has been manufactured on a technically important scale. There have been great technological improvements, and the quality of the product has been greatly enhanced, as is exemplified by modern motor tyres which have a far longer service life than their early forerunners. Rubber originally came from South America and was used by the native population for making balls which had exceptional bouncing properties. It was in this form that Columbus got to know rubber. Three centuries elapsed before the material was brought into commercial use in Europe, and at first it was used not for its elastic properties but to rub out lead pencil marks.

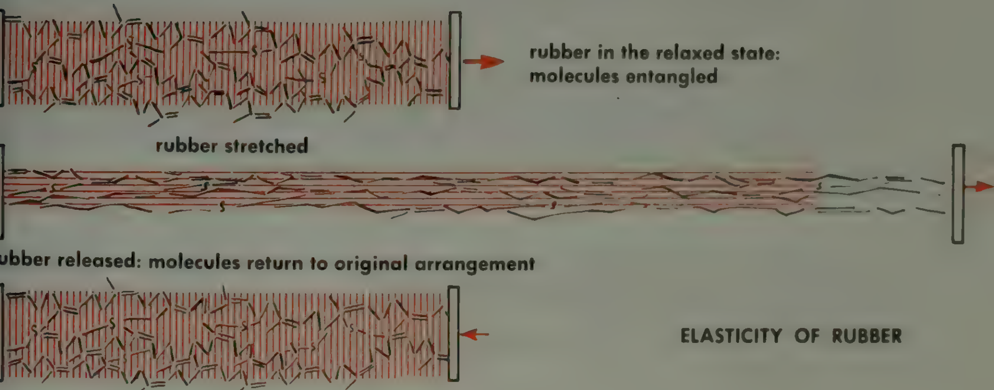
The world's annual output of natural rubber (plantation rubber) is now about 2 million tons. In addition, various kinds of synthetic rubber is being used on an ever increasing scale. See page 56.

1. Tires in U.S.A.



rubber molecule





ELASTICITY OF RUBBER

SYNTHETIC RUBBER (BUNA)

The name "Buna" is applied to a group of synthetic rubbers first developed in Germany and is produced by a process of polymerisation from *bu*tadiene with sodium (*na*trium) as a catalyst.[1] The process used to be carried out at a temperature of about +50° C and yielded "lettered" Buna rubbers such as, for example, Buna S (butadiene styrene rubber). Nowadays copolymerisation of butadiene and styrene is mostly done in aqueous phase. With the newer activators it is possible to carry out this process at about +5° C, whereby the present form known as "cold rubber" is obtained. By appropriate variation of the monomers, their proportions (chiefly about 75% butadiene and 25% styrene) and the polymerisation conditions, a number of different types of Buna rubber are obtained, and this range of types can be further extended by various methods of processing and by using various admixtures. Latterly, with aid of so-called Ziegler catalysts, a product bearing a closer resemblance to natural rubber can also be produced from butadiene or isoprene, e.g., Buna CB (poly-cis-butadiene).

In the emulsion copolymerisation process carried out at +5° C (as referred to above), the hydrocarbons to be polymerised (e.g., butadiene and styrene) are in emulsion and contain a constituent of the activator system dissolved in them. The second part of the activator system is present in the aqueous phase (the watery medium of the emulsion). The combined activator system initiates the process of polymerisation. The molecule size of the polymer obtained can be regulated by certain added substances. The macromolecules (giant molecules of very great length) formed in this way have a filamentary structure with branches, so-called side chains. The polymerisation of the monomers is stopped after about 60% of these substances have reacted. The resultant product at this stage is a latex rather like the latex of natural rubber (see page 54). The unreacted monomers are removed from this latex, and stabilisers are added to it, whereafter the latex is coagulated by the addition of acids and salts. The solid matter obtained in this way is washed and dried in several stages.

For processing Buna into rubber goods, it is treated in masticating machines or on mixing rollers, various substances being added whereby the workability of the rubber and/or the properties of the vulcanisates are controlled. Such admixtures are, for example: oils, paraffin, fatty acids, tars, bitumen, carbon black, zinc oxide, chalk, silica, kaolin, finely divided organic and inorganic substances. For vulcanisation, which is usually carried out under pressure at approximately 150° C, the mixture moreover has sulphur and vulcanisation accelerator (e.g., mercapto benzothiazole) added to it.

In the process of vulcanisation the filamentary molecules become interlinked into a three-dimensional network, the "links" between the molecules being formed by sulphur. The process is known as cross-linking. As a result, the rubber largely loses its plastic properties and, instead, acquires a high degree of elasticity and other properties associated with manufactured rubber (e.g., wear resistance). Buna is used for making motor tyres,[2] rubber conveyor belts, and many other technical products.

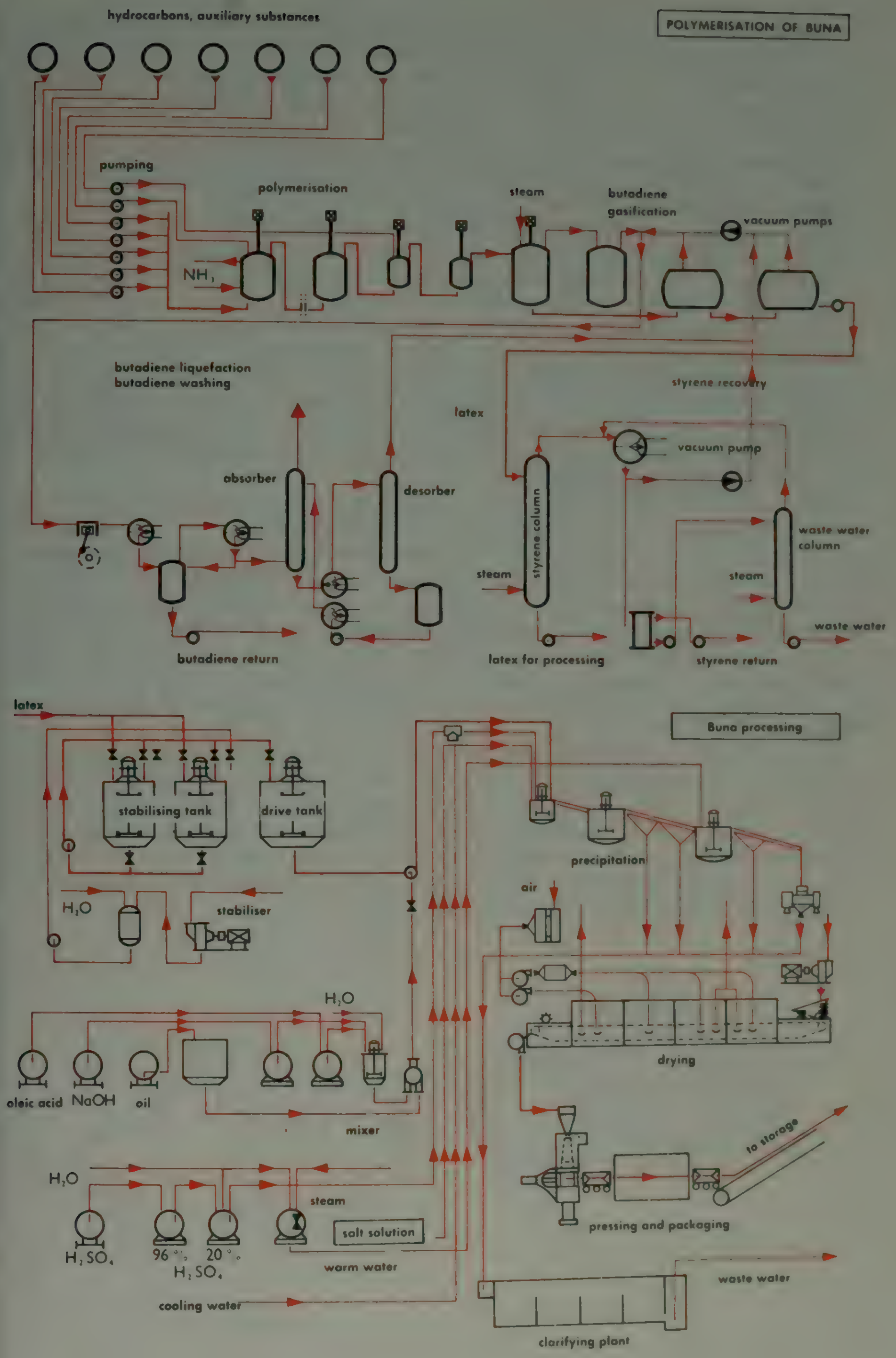
There are many other synthetic rubbers besides the Buna rubbers. For example, Perbunan, Hycar, Chemigum and Butaprene are nitrile-butadiene rubbers of the oil-resistant type. During the Second World War a synthetic rubber named GR-S (Government Rubber, styrene type) was developed in the United States and extensively used. Other rubbers are butyl (GR-I) and neoprene (GR-M).

1. Polymerisation: a reaction involving a successive addition of a large number of relatively small molecules (monomers) to form the final compound or polymer.

Polymer: giant molecule (macromolecule) formed when thousands of the original molecules have been linked together end to end.

Copolymer: a mixed polymer; a giant molecule formed when two or more unlike monomers are polymerised together.

2. Tires in U.S.A.


POLYMERISATION OF BUNA
hydrocarbons, auxiliary substances
pumping
polymerisation
NH_3
steam
butadiene gasification
vacuum pumps
butadiene liquefaction
butadiene washing
latex
styrene recovery
absorber
desorber
vacuum pump
styrene column
waste water column
steam
steam
waste water
butadiene return
latex for processing
styrene return
latex
Buna processing
stabilising tank
drive tank
precipitation
air
H_2O
stabiliser
H_2O
drying
oleic acid
NaOH
oil
mixer
to storage
H_2O
steam
pressing and packaging
salt solution
H_2SO_4
96 %
20 %
H_2SO_4
warm water
waste water
cooling water
clarifying plant

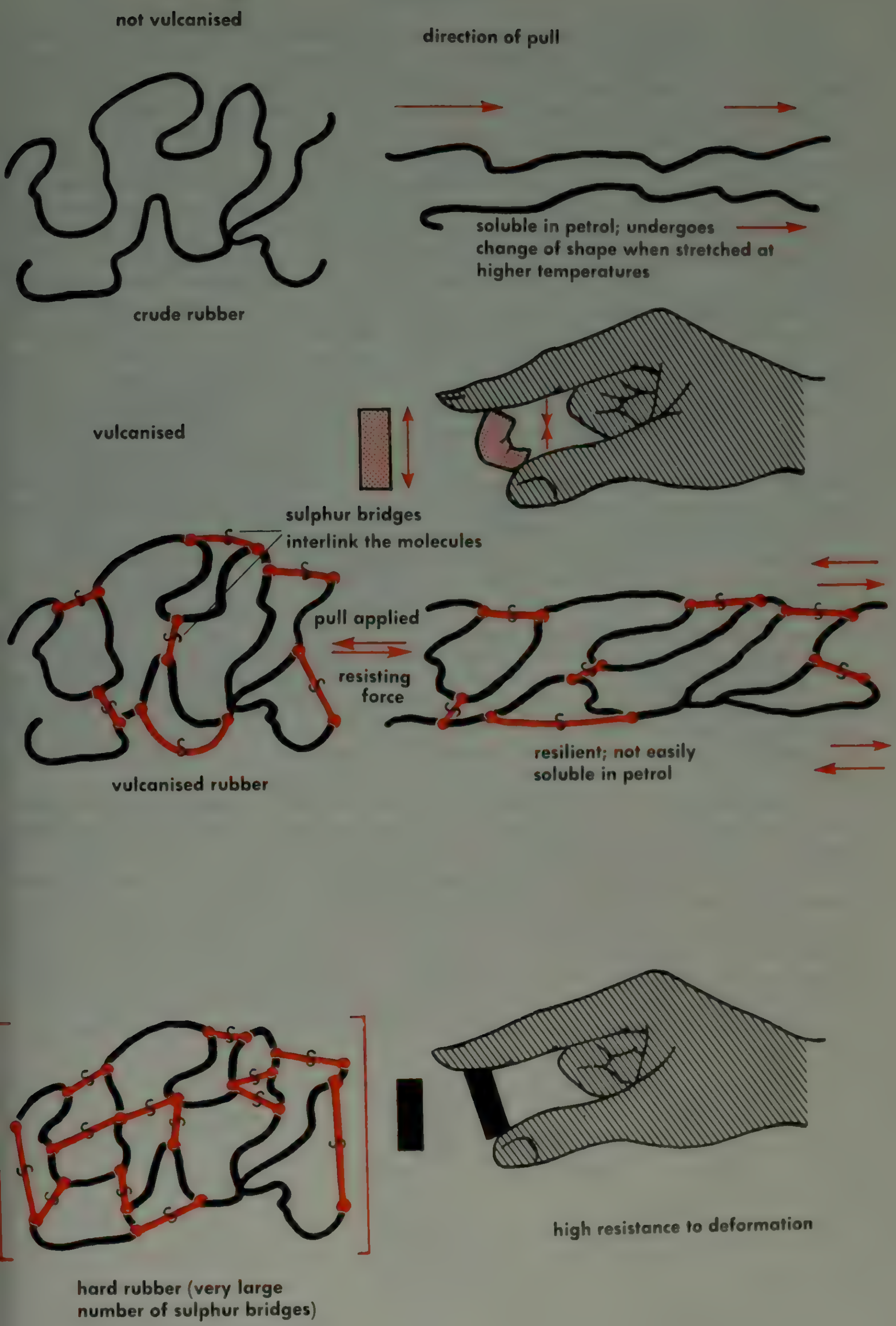
Vulcanisation or curing is a chemical reaction whereby the filamentary molecules of rubber are interlinked into a three-dimensional network, this being usually achieved with the aid of sulphur. Sometimes peroxides are used for the purpose, however. It was Goodyear who, in 1839, first masticated crude rubber with sulphur and heated the mixture to 130° C. After undergoing this treatment, the rubber was no longer plastically deformable but, instead, acquired a high degree of resilience which was retained over a wise range of temperatures. The solubility of crude rubber in petrol[1] is greater than that of vulcanised rubber.

Because of the wide range of products for which rubber is used (e.g., motor tyres,[2] tubing, seals, footwear, gloves, etc.), it is necessary to incorporate other admixtures besides sulphur into the crude rubber. Various substances are mixed into the rubber in masticating machines or on roll mills, e.g., carbon black (for high abrasion resistance), silica, chalk, asbestos (more particularly for brake linings), oils (for better workability of the mixture), paraffin (for better resistance to light), antioxidants (usually : aromatic amines or phenol derivatives), activators (usually zinc oxide), and various organic and inorganic colouring substances. In order to speed up the vulcanisation process and to improve the properties of the vulcanisates, various accelerators are added, e.g., dithio carbamic acid derivatives, mercapto benzothiazole derivatives, diphenylguanidine, etc.

Vulcanisation is carried out under pressure in moulds at temperatures around 150° C and takes from a few minutes to several hours, depending on the vulcanisation temperature and the size of the rubber article concerned. Vulcanising an ordinary motor tyre takes about half an hour. By using special combinations of accelerators it is also possible to perform the vulcanisation process at ordinary room temperature. Some rubber mixtures are manufactured into various special sections (tubes, sealing gaskets for car windows, etc.) by extrusion. Such extruded articles are vulcanised under pressure in vulcanising vessels. Other mixtures are processed by calendering, i.e., the rubber is pressed between rolls to form sheets of predetermined size and thickness.

Sponge rubber is usually produced from latex, which is foamed by various methods and then vulcanised. Certain rubber mixtures can be bonded to metals so as to establish a permanent connection. Soft rubber contains about 1.5–5.5 and hard rubber contains about 15–30% sulphur. In cases where rubber goods have to fulfil special requirements—e.g., high resistance to swelling in organic solvents, to the action of light, or to high temperatures—it may be necessary to use certain synthetic rubbers, such as Perbunan or butyl rubber (see page 56).

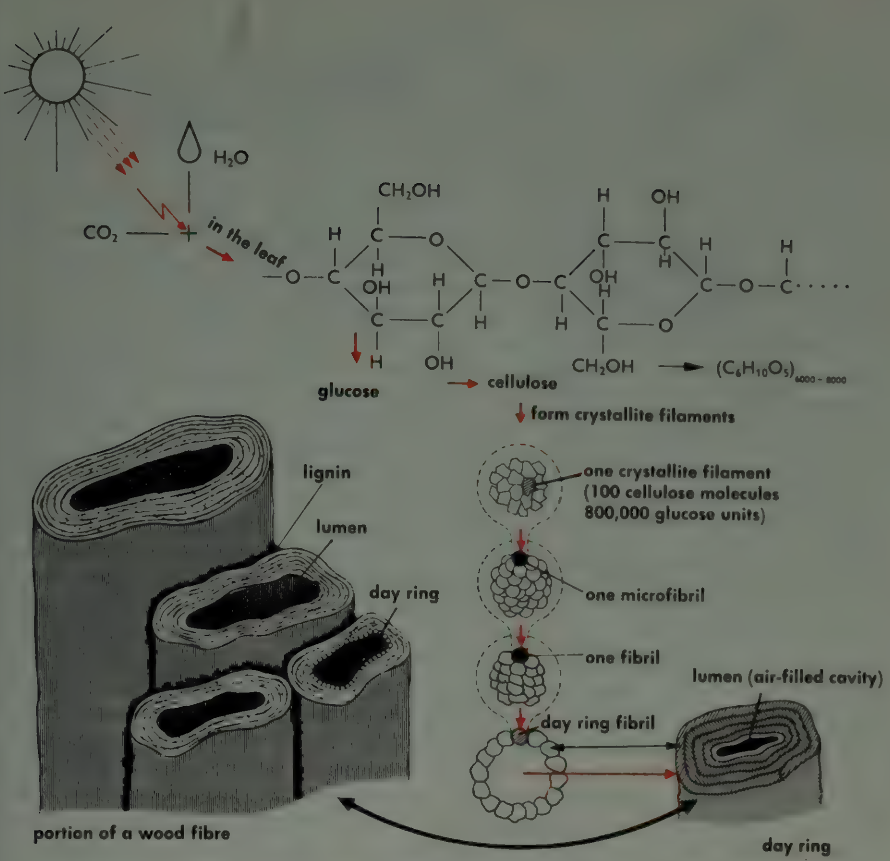
1. Gasoline in U.S.A.
2. Tires.


not vulcanised
direction of pull
soluble in petrol; undergoes change of shape when stretched at higher temperatures
crude rubber
vulcanised
sulphur bridges interlink the molecules
pull applied
resisting force
resilient; not easily soluble in petrol
vulcanised rubber
high resistance to deformation
hard rubber (very large number of sulphur bridges)

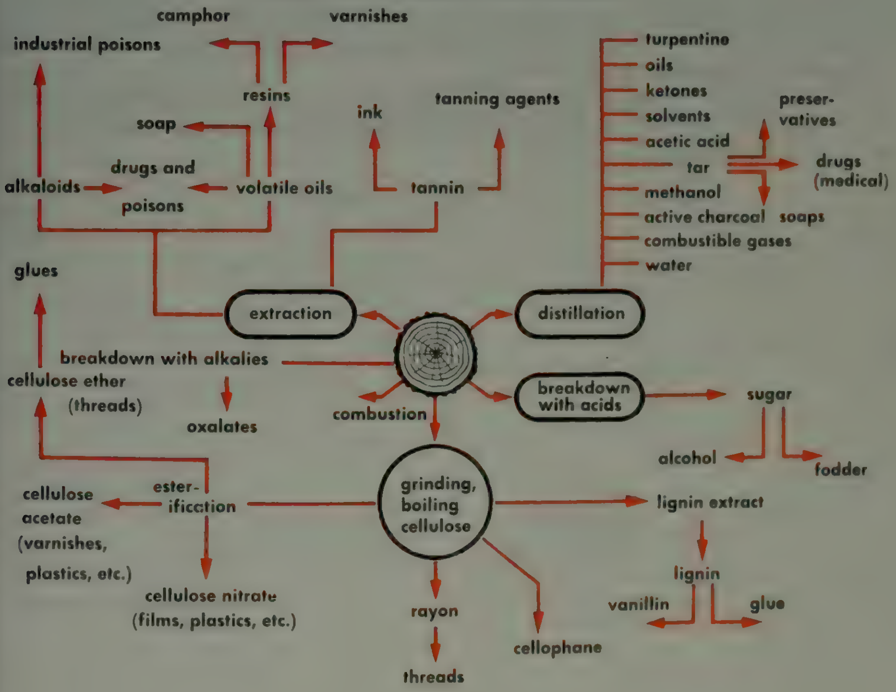
Along with stone, timber is one of mankind's oldest building materials. It is still a very important commercial commodity. At the present time, however, timber is being felled at a far greater rate than it can grow, so that it is bound to become increasingly scarce. Wood cannot be produced synthetically in the laboratory. With the aid of the sun's light and heat, plants convert carbon dioxide from the atmosphere and moisture from the soil into glucose, some of which they store up in the form of starch, while another portion is converted into cellulose, from which elongated cells are formed. In this way. complex bundles of tubes (called fibres) are formed through which water and other substances essential to plant life are conveyed and which also serve to give mechanical strength to the plant stems. The individual glucose molecules which are synthesised in the leaves undergo a process of polymerisation whereby they form chain molecules containing something like six to eight thousand glucose monomers. About a hundred of such chain molecules form a crystallite filament (about 6–8 millionths of a millimetre thick); between ten and a hundred of these filaments in turn form a microfibril (about 20–80 millionths of a millimetre thick). A bundle comprising 10–100 microfibrils form a fibril—the smallest unit which is just visible under an ordinary microscope when a wood fibre is finely ground up. The fibrils are arranged side by side, in so-called day rings, and are helically twisted. The individual fibrils are bonded together by a substance called lignin (a compound similar to cellulose). The day rings are shaped like hollow cylinders, a large number of which are disposed one around the other; at the centre of a composite tube of this kind is an air-filled cavity (lumen), which is 2-3 ten thousandths of a millimetre thick. Finally a number of such tubes firmly bonded together form a wood fibre, which is usually about $\frac{1}{30}$ mm (0.0013 inch) thick and is interpenetrated by tiny passages ranging in diameter from $\frac{1}{10}$ to $\frac{1}{100{,}000}$ mm.

Green timber contains 40–60%, seasoned timber contains 10–20% moisture. Completely dehydrated wood consists of 40–50% cellulose and 20–25% lignin, the remaining constituents being minerals, tanning agents, fats, oils, resins, carbohydrates, and hemicelluloses. Wood is a particularly useful material on account of low thermal conductivity, its high absorptive capacity for radiation (heat rays, etc.), and its strength.

A main distinction is made between softwoods (mainly from coniferous trees: pine, deal, fir, etc.; also poplar, willow, etc.) and hardwoods (oak, beech, etc.). The seasoning of timber is of paramount importance for the majority of purposes for which it is used. In green timber there is a large amount of free water present within the cells. It is the drying-out of the free water and a certain amount of moisture from the cell walls that constitutes seasoning. There are two main methods of seasoning: air seasoning (boards are stacked up in such a way as to obtain a good air flow throughout the stack and thereby evaporate the moisture) and kiln seasoning (this is done in a kiln, which is a brick-built chamber equipped with heating pipes, fans to keep the air circulating, and a number of jets for the introduction of steam).



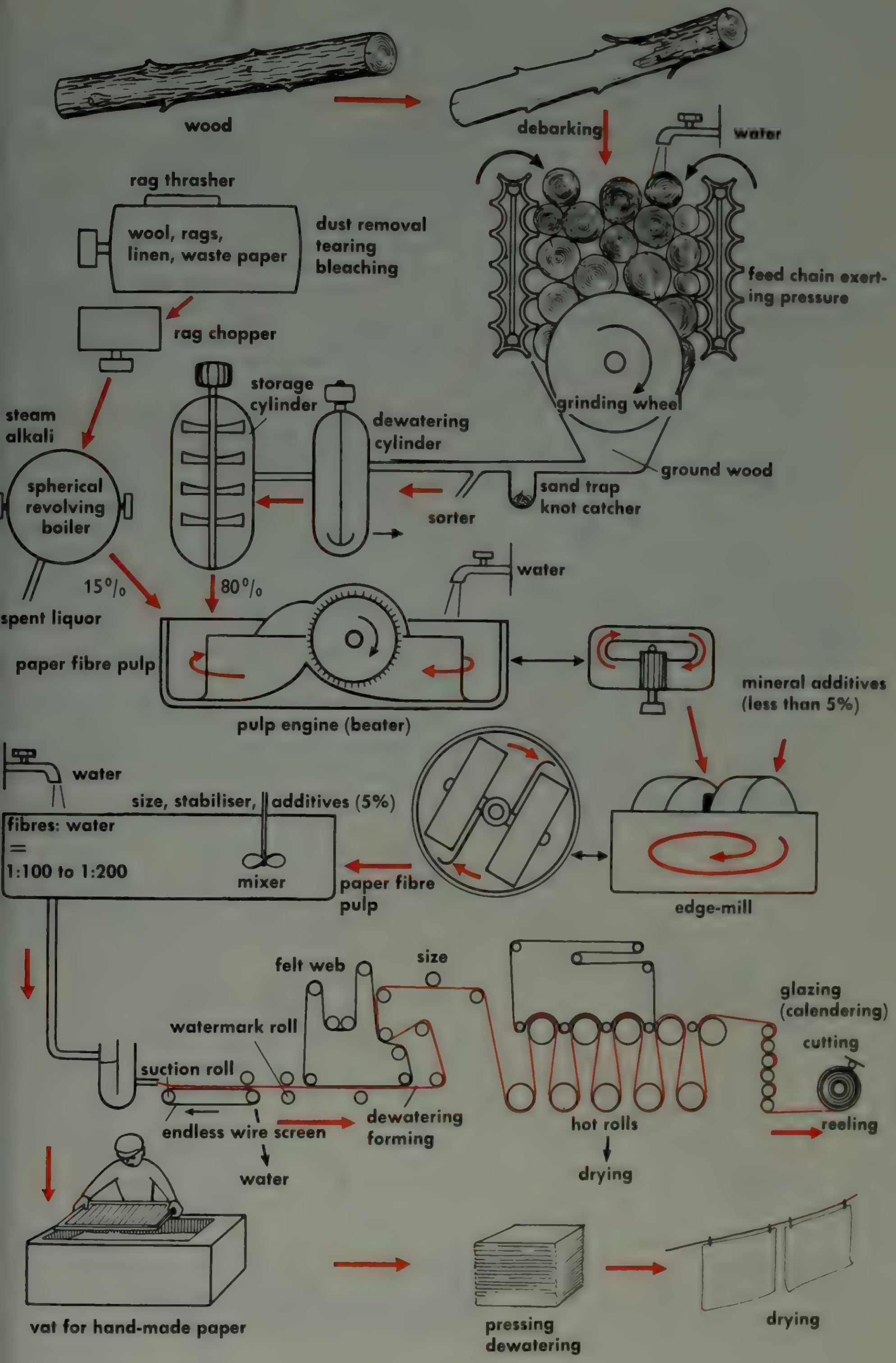
WOOD AS A CHEMICAL RAW MATERIAL



Paper is the general name for the substance used for writing or printing upon or for packaging and wrapping, as well as for many special purposes. It derives its name from the papyrus plant from which the ancient Egyptians produced a paper-like material for writing on. Modern paper is manufactured from a mixture of various fibres, chiefly of vegetable origin, which are mixed with a large quantity of water and shredded very fine. The mixture is treated with size (a glue-type admixture which makes it water resistant), and a filler, such as clay or chalk, may be added to give some special property to the paper. The result of this preparation is called "stuff". It is poured out on wire screens on which the water is extracted.

The fibrous ray material that is used in present-day paper making is wood from coniferous trees (pine, fir, spruce), deciduous trees (beech, poplar, birch, chestnut, eucalyptus), grasses (straw from rye, wheat, oats, barley, rice), alfa grass, bamboo and cotton. Biochemical or chemical and mechanical dissolving and comminution processes are employed. The principal raw material for newsprint—wood pulp—is obtained by grinding up whole logs with the aid of special grindstones and adding a considerable quantity of water. Small quantities of other fibrous materials such as hemp, linen, wool, asbestos, slag wool, glass fibres and synthetic fibres are sometimes added to the vegetable fibres. The fibre mixture is chemically bleached, further broken down in beaters, mixed with fillers and more water (about 100 to 200 parts of water to one part of fibre) and constantly stirred, in conjunction with the addition of colophony soap, animal glues, starch paste, casein or synthetic resin glues and certain other substances. These admixtures ensure that, despite the great dilution with water, the size is deposited almost entirely on the fibrous constituents. The paper pulp is poured out on screens (in paper machines these may be endless wire screens), dewatered, pressed, watermarked, and (in some cases) it undergoes subsequent sizing treatment, drying, calendering (whereby a smooth finish is obtained by passing the paper through highly polished rolls), and cutting to the required size. The composition of the paper varies according to the purpose for which it is used. Roofing felt, which is really an asphalted cardboard, is made, for example, with peat fibre in addition to wood pulp; document papers consist of fibre mixtures containing linen and hemp; special papers contain certain textile and animal wool fibres; newsprint comprises 80% wood pulp (ground wood), 15% cellulose, 5% glue and filler materials. The method of processing also plays an important part (surface treatment, very fine grinding for transparent paper, etc.).

Paper manufacturing machines are among the longest and most delicately adjusted pieces of machinery used in any industry. They are so long because the paper web has to be dried, smoothed and cooled.


wood
debarking
water
rag thrasher
wool, rags,
linen, waste paper
dust removal
tearing
bleaching
feed chain exert-
ing pressure
rag chopper
storage
cylinder
steam
alkali
dewatering
cylinder
grinding wheel
ground wood
spherical
revolving
boiler
sand trap
knot catcher
sorter
15%
80%
water
spent liquor
paper fibre pulp
pulp engine (beater)
mineral additives
(less than 5%)
water
size, stabiliser,
additives (5%)
fibres: water
=
1:100 to 1:200
mixer
paper fibre
pulp
edge-mill
felt web
size
glazing
(calendering)
watermark roll
cutting
suction roll
endless wire screen
dewatering
forming
hot rolls
reeling
water
drying
vat for hand-made paper
pressing
dewatering
drying

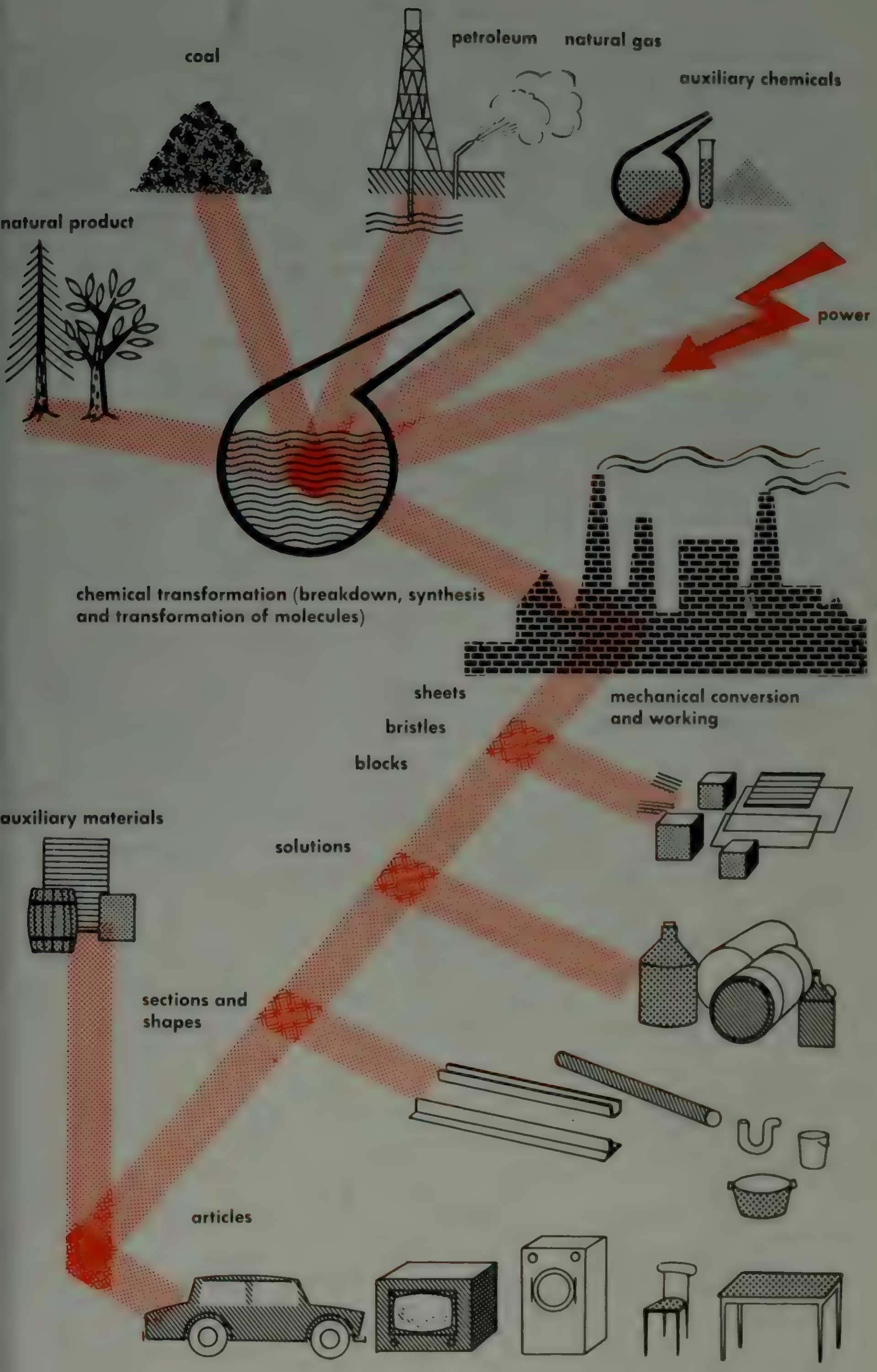
Plastics are materials consisting of macromolecular organic compounds which are produced from animal or vegetable matter or from raw materials derived therefrom. The general name "plastics" is based on the fact that these materials are of plastic consistency at some stage of their manufacture.

Man-made fibres and other synthetic substances of this type are not, however, included in the definition of "plastics" in the more restricted sense of the word.

Plastics are produced from a variety of raw materials: wood, coal, petroleum, plants, vegetable juices, proteins. In some respects they are superior to comparable natural products because they are produced under scientifically controlled conditions and can be given specific properties not found in nature.

All plastics consist of very large organic molecules, so-called macromolecules, which usually contain from 500 to 10,000 equal "structural units" chemically bounded together. Hard, inelastic (rigid) plastics consist mainly of spherical, highly cross-linked macromolecules; plastics possessing rubber-like elastic properties consist chiefly of thread-like macromolecules.

Some plastics such as celluloid, nitrocellulose and chlorinated rubber are derived from naturally pre-formed macromolecules; such plastics, are, in effect, chemically modified or improved natural substances. The fully synthetic plastics such as synthetic butadiene rubber (Buna rubber), polystyrene, acrylo-nitrile, polyvinyl chloride, etc. must, however, be built up from small molecules by processes called polymerisation and polycondensation. Some of the good properties of plastics are: low specific gravity, easy deformability, resistance to many corrosive chemicals, non-toxicity, excellent electrical properties, etc. Plastics are used in electrical engineering (insulation for cables, wires, equipment), in the clothing industry, in the manufacture of sheeting, films, threads, brushes, adhesives, artificial leather, linoleum, paints, etc. These materials have become indispensable for industrial and domestic purposes, as they are increasingly replacing metals and other materials which are not available in sufficient quantity or at a low enough price, and they are opening up entirely new possibilities in mechanical engineering, in vehicle construction, in the furniture industry, and in the packaging of perishable goods. The first "plastic" was produced about a hundred years ago—at the present time these synthetic materials are still in full process of development.

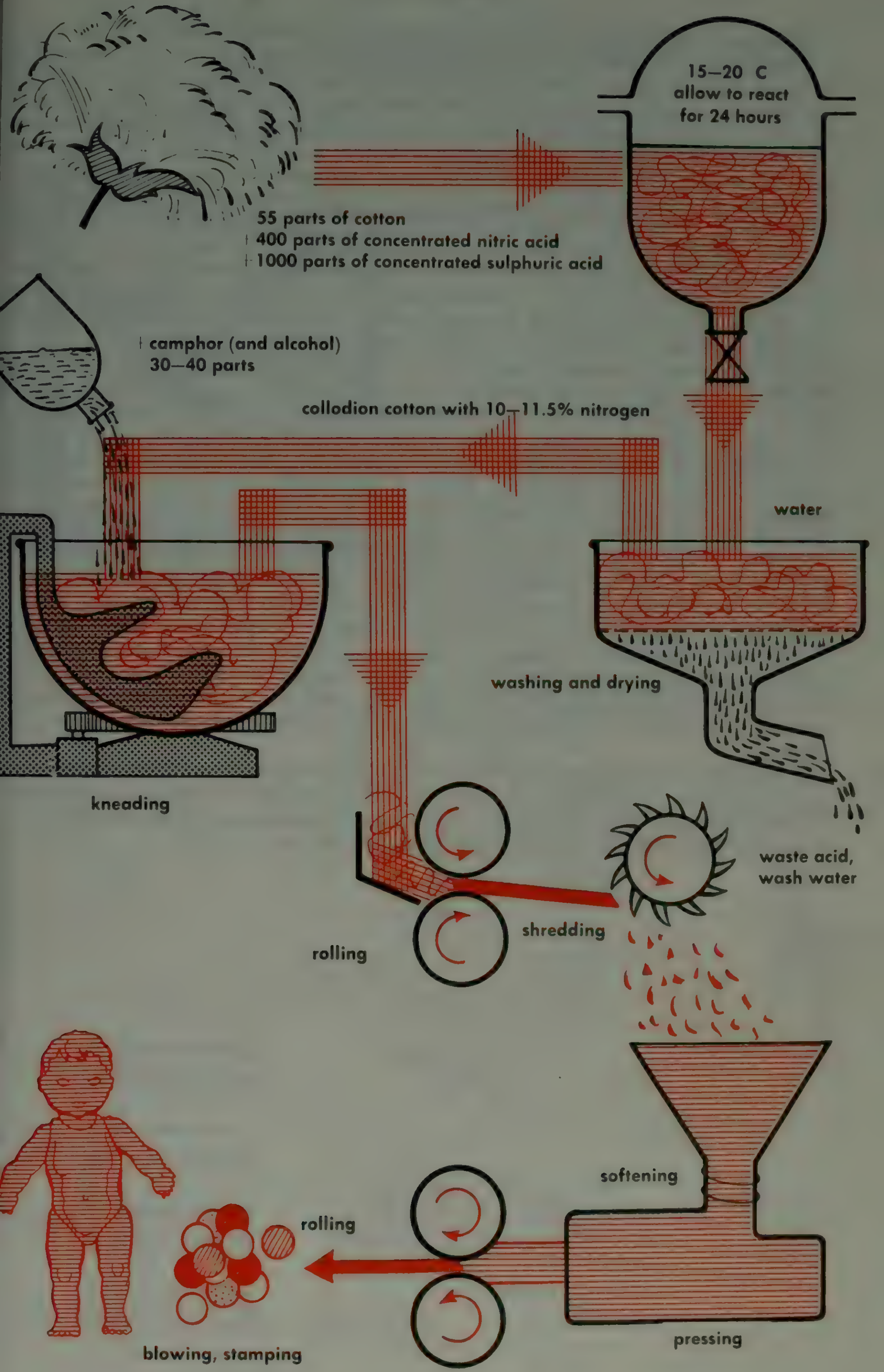

coal
petroleum
natural gas
auxiliary chemicals
natural product
power
chemical transformation (breakdown, synthesis and transformation of molecules)
sheets
bristles
blocks
mechanical conversion and working
auxiliary materials
solutions
sections and shapes
articles

CELLULOID

Celluloid is one of the oldest plastics. It was first produced in 1869. It consists of slightly nitrated cellulose (nitrocellulose with 10–11.5% nitrogen; chemical name: dinitrocellulose) and camphor as a softener.

Nitrocellulose is made from cotton by nitration with nitric acid and sulphuric acid. The dried nitrocellulose is thoroughly kneaded with a solution of camphor in alcohol: 30–40 parts of camphor to 100 parts of nitrocellulose. The kneaded mixture is now rolled into plates or is shredded, dried, and then hot-moulded. Celluloid is hard and resilient, colourless, and used to be very extensively employed in the manufacture of photographic films. It dissolves very well in acetone and softens at temperatures above 80° C. The one great disadvantage of this versatile plastic is its inflammability, and for this reason its use for films was soon abandoned.

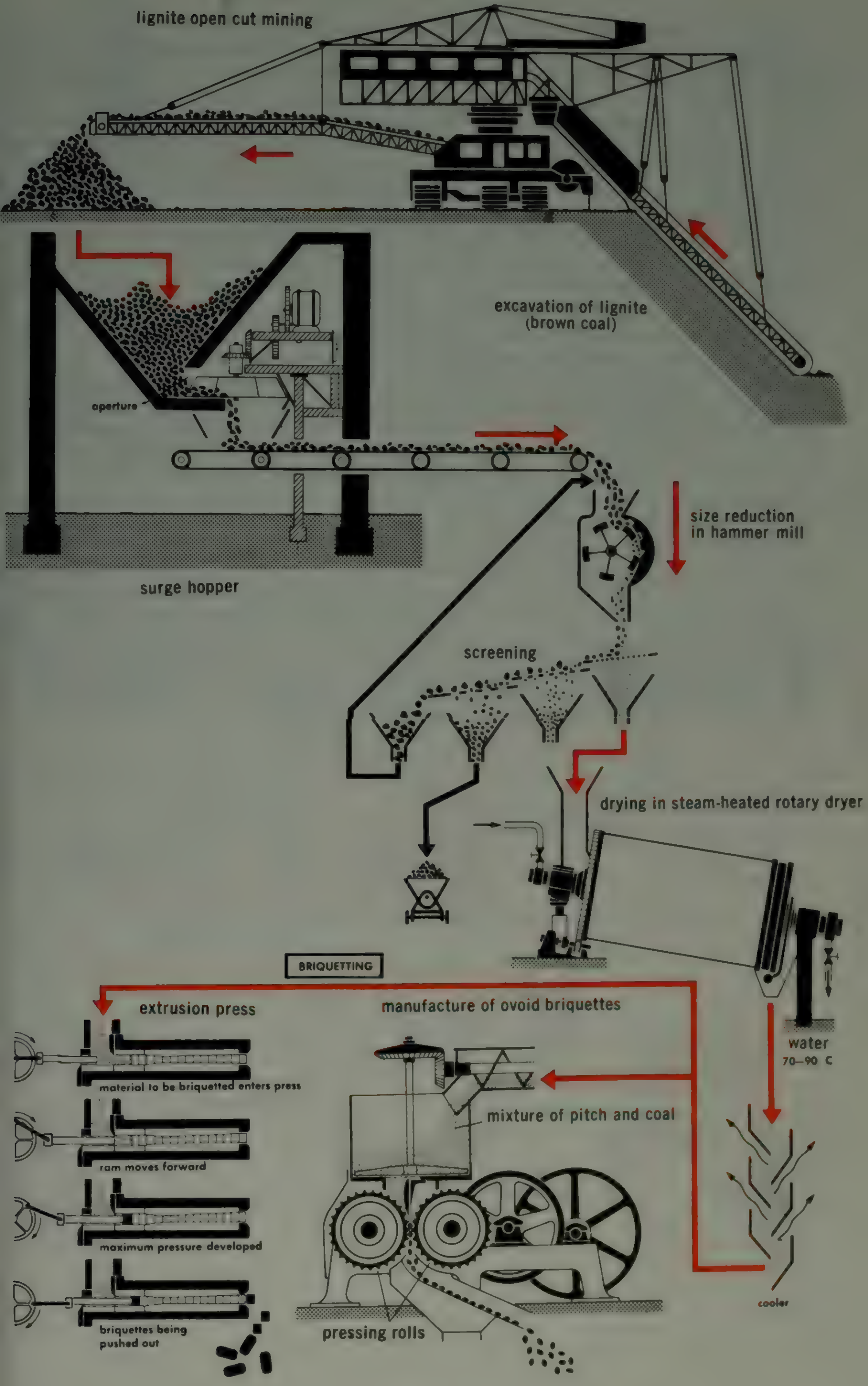
As celluloid can be easily worked, it is still extensively used for manufacture of a wide variety of objects, e.g., toys, dolls, ping-pong balls, brushes, combs, knobs, and many others. As such it serves as a substitute for such natural materials as ivory, horn and tortoiseshell. The addition of substances called stabilisers to the mixture of nitrocellulose and camphor gives the final product better fastness to light and helps it retain its resilience for a longer time. Celluloid achieved its greatest popularity in the early days of the film industry. At the present time it is being superseded more and more by materials which are non-inflammable or harder or with better elastic properties. Such materials are usually not so universally applicable as celluloid, but they are better suited to specific purposes. Nevertheless the celluloid manufacturing industry still consumes over two-thirds of the present-day output of synthetic camphor.


15–20 C
allow to react
for 24 hours
55 parts of cotton
+ 400 parts of concentrated nitric acid
+ 1000 parts of concentrated sulphuric acid
+ camphor (and alcohol)
30–40 parts
collodion cotton with 10–11.5% nitrogen
water
washing and drying
kneading
waste acid,
wash water
shredding
rolling
softening
rolling
pressing
blowing, stamping

Brown coal (or lignite) may contain as much as 60% moisture and often has a low calorific value. To obtain a higher-grade product, brown coals can be briquetted, i.e., squeezed into compact lumps by a suitable pressure-moulding process. To do this, the brown coal is pulverised, dried to a water content of 16-18%, and then pressed to briquettes. As a result, the calorific value is more than doubled in comparison with that of raw brown coal. The brown coal, which is usually mined by open-cast methods, is dug out of the ground by an excavator and conveyed to a receiving hopper. From here it is passed through crushing and pulverising machinery (e.g., hammer mills). The pulverised brown coal is screened, the oversize material being returned to the pulverising mills or sold as "screened rough coal". The coal that passes the screens still contains a high percentage of moisture. It is passed to dryers whereby its moisture content is reduced to 17–20%. A modern rotary dryer is an inclined rotating drum, heated by steam, into which the wet material is fed at the higher end and from which the dried material is discharged at the lower end.

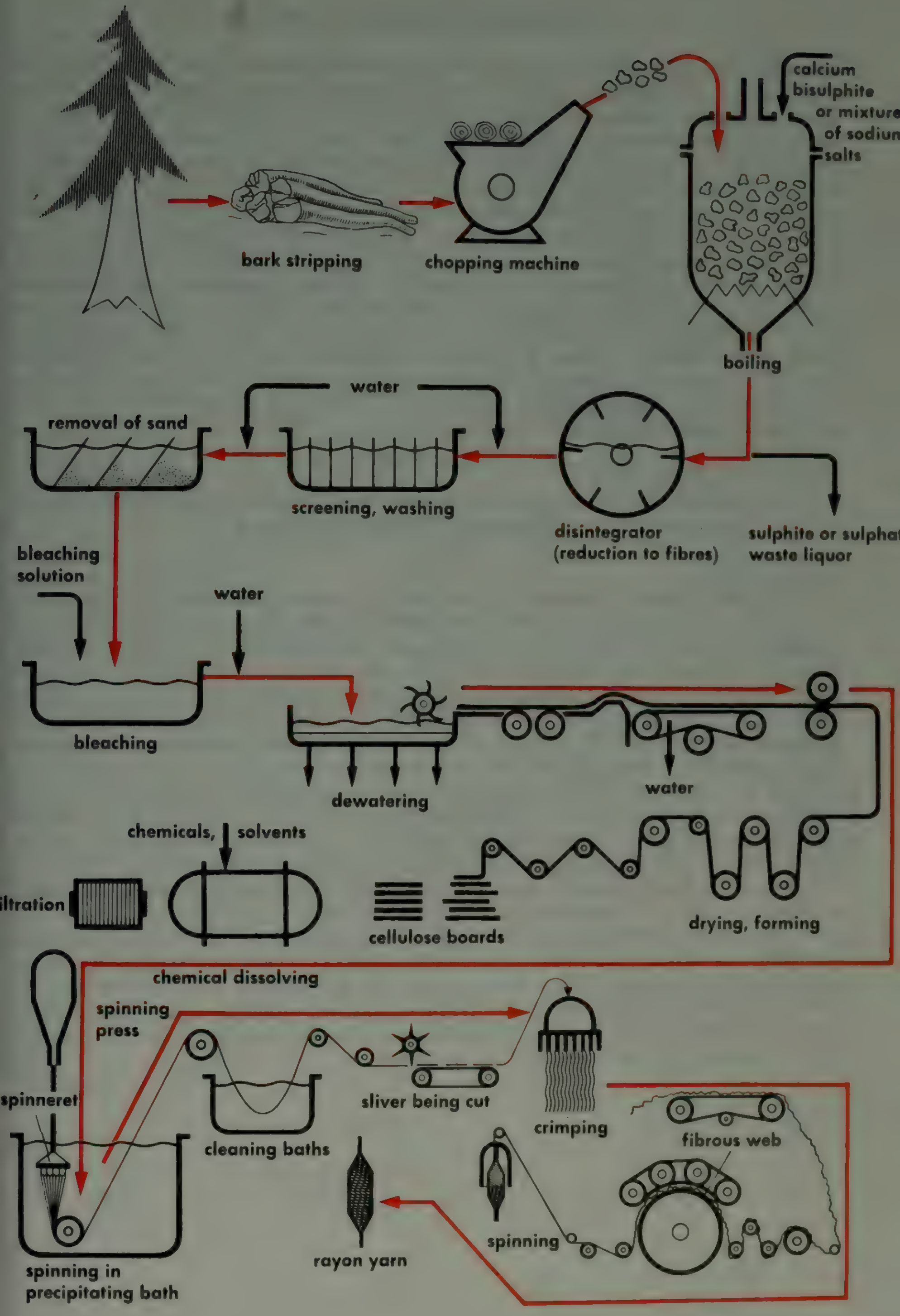
The coal, which is heated to 80°–90° C in the drying process, is then cooled to 40°–50° C in a special cooler provided with baffle plates. It now contains about 15–18% water and is fed to the briquetting presses. These may operate on the extrusion principle: a certain quantity of the dried pulverised brown coal is thrust forward and compressed by each stroke of a reciprocating ram which forces it through a tapered opening. The ram has a stroke of about 200 mm (8 in.) and develops a pressure of 700–1000 kg/cm^2 (10,000–14,000 lb./in.^{2}). The process generates a great deal of heat, and the freshly formed briquettes are given time to cool before they are loaded into vehicles for transport to the consumers. A large press may have an output of 250 tons of briquettes per day. The briquettes are formed by pressure alone, no binder being employed. Another type of briquetting press is the ring roll press, which attains similar outputs. If the dried screened coal has a low water content (7–10% moisture), it is necessary to use a pressure of about 2000 atm. (approx. 30,000 lb./in.^{2}) in order to obtain firm briquettes. The dust produced in the process is extracted by suction equipment, and, where possible, used in pulverised-coal firing systems for boilers. Ovoid (egg-shaped) briquettes are produced in special roll presses.

Briquetting is not confined to brown coal. Ordinary coal dust and fines are moulded into briquettes with the aid of pitch and coal tar as binders. Coal briquettes are used for domestic and industrial purposes. Brown coal briquettes are chiefly employed as household fuel in the neighbourhood of their manufacture.


lignite open cut mining
excavation of lignite
(brown coal)
aperture
surge hopper
size reduction
in hammer mill
screening
drying in steam-heated rotary dryer
BRIQUETTING
extrusion press
material to be briquetted enters press
ram moves forward
maximum pressure developed
briquettes being
pushed out
manufacture of ovoid briquettes
mixture of pitch and coal
pressing rolls
water
70—90 C
cooler

CELLULOSE

Cellulose (wood pulp) is prepared from wood by a process of dissolving and chemical softening. It serves as a raw material for the manufacture of paper, nitrocellulose, explosives, varnishes, wood sugar (xylose), collodion, celluloid, and cell wool. Depending on the solvent solution used, the following pulping processes are distinguished: sulphite process; soda process; sulphate process. The pulp is produced from pine, fir, spruce and beech wood and also from other cellulose-containing vegetable matter, such as sugar-cane waste, straw, reeds, maize and sunflower stalks. These materials are subjected to a preliminary cleaning treatment, chopped up by mechanical processes, and then boiled in large tanks with hot solutions (calcium bisulphite solution or a mixture of caustic soda, sodium sulphide, sodium sulphate, sodium carbonate). The lignin which causes the cellulose fibres to adhere together is dissolved by this treatment and is, in part, chemically decomposed, leaving a soft pulp which consists of cellulose (see timber, page 60). The raw pulp thus obtained is reduced by mechanical processes, washed, bleached and again carefully cleaned. Finally, the pulp is dewatered and formed into strips or boards. To make 1000 lb. of cellulose requires about 3 $yd.^3$ of wood, 500 lb. of coal, nearly 100 lb. of sulphur (if the sulphite process), and 90 kWh of electric energy. About 10 million tons of cellulose are produced each year. As already stated, this material is used for the manufacture of paper (see page 62), nitrocellulose, explosives (see page 172), varnishes, wood sugar, celluloid (see page 66), cellite, cellon collodion, cellulose ether, cellulose ester, cellophane, and other products. Man-made fibres (see page 72) and cell wool are also manufactured from cellulose. For the manufacture of cell wool, cellulose is chemically dissolved and spun into threads in precipitating baths. The skein of threads is washed, cleaned, and cut up into 1–6 in. long pieces. These staple fibres are curled, prepared, dried, opened up, and worked into a spun yarn. The lustre, curliness, fineness, and length of cut are made similar to those of natural wool or cotton. The properties of cell wool have been steadily improved in recent years, with the result that it has attained and even surpassed the quality of cotton. Cell wool is used by itself and also in a mixture with natural wool. Most of the cellulose produced is consumed by the textile industry and the paper industry.



bark stripping
chopping machine
calcium bisulphite or mixture of sodium salts
boiling
water
removal of sand
screening, washing
disintegrator (reduction to fibres)
sulphite or sulphat waste liquor
bleaching solution
water
bleaching
dewatering
water
chemicals, solvents
iltration
cellulose boards
drying, forming
chemical dissolving
spinning press
spinneret
cleaning baths
sliver being cut
crimping
fibrous web
spinning
rayon yarn
spinning in precipitating bath

As long ago as 1664 the English scientist Robert Hooke occupied himself with the question how silk filament might be produced without the intervention of silkworms. His experiments yielded no useful results, however. To produce a fibre from any particular raw material, three problems have to be solved. In the first place it is necessary to have a raw material which can be melted or dissolved; secondly, it must be possible to extrude the melted material or the solution from very fine holes or to form it into coagulable threads in some other way; thirdly, the threads must be very flexible and yet very strong, they must not dissolve in water, they must have some degree of heat resistance, and they must possess the necessary textile properties.

The vast field of man-made fibres (synthetic fibres) is subdivided into two main groups: fibres made from cellulose, and true synthetic fibres (see page 78). In both groups filaments and spinning fibres are produced. The former are thousands of yards long (virtually "endless"), whereas the latter are only a few inches in length; they are cut or torn to the length of the natural fibres with which they are often used in combination. Cellulose-based fibres are at present still a commercially far more important commodity than true synthetic fibres.

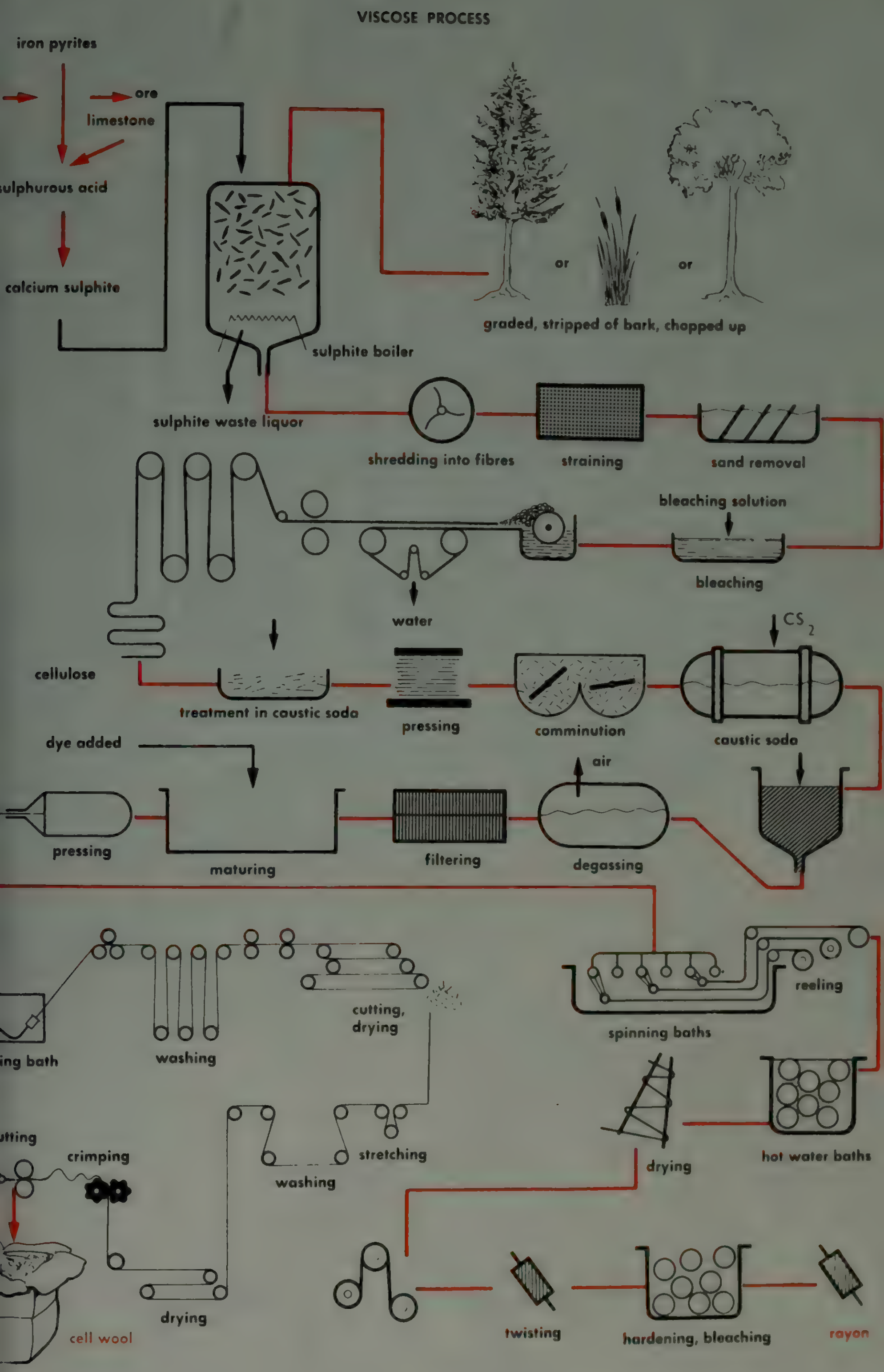
Cellulosic man-made fibres

At present these fibres are manufactured by three different processes: the viscose process, the cuprammonium process, and the acetate process.

Viscose process

The initial material is cellulose, produced by chemical processing of wood, straw and various other vegetable substances (see page 70). If cellulose is treated with caustic soda, the low-molecular components are dissolved in it, while the high-molecular components (which are the useful ones for the production of man-made fibres) form a salt which can react with hydrogen sulphide at 20°–25° C to produce cellulose xanthate. This thick yellow substance dissolves in 7% caustic soda (sodium hydroxide solution) at 15°–17° C to form a viscous solution (viscose), which is then filtered and freed from gas (by vacuum treatment) because even the tiniest gas bubbles would cause the filament to snap during spinning. The spinning solution must mature at 12°–15° C for 2–4 days before spinning. It is then pumped through spinning nozzles into spinning baths. Each nozzle is provided with a large number (up to 150) very fine holes called spinnerets ($\frac{1}{20}$–$\frac{1}{10}$ mm = 0.002–0.004 in. diameter) and are made of a chemically resistant metal such as tantalum. The spinning bath contains a solution of sulphuric acid and salts in water. In an acid solution of this kind the cellulose xanthate is decomposed into cellulose in the form of filaments emerging from the nozzle. These are then passed through various cleaning, bleaching, hardening, and refining baths. After drying, 10–20 filaments are spun together to form the rayon yarn. Depending on the pre-treatment and after-treatment to which it has been subject, the rayon is colourless or already spun-dyed (a very wash-fast dyeing process), lustrous or dull, fine or coarse.

(Continued)



VISCOSE PROCESS
iron pyrites
ore
limestone
sulphurous acid
calcium sulphite
or
or
graded, stripped of bark, chopped up
sulphite boiler
sulphite waste liquor
shredding into fibres
straining
sand removal
bleaching solution
bleaching
water
cellulose
treatment in caustic soda
pressing
comminution
CS_2
caustic soda
dye added
air
pressing
maturing
filtering
degassing
reeling
spinning baths
cutting, drying
washing
ing bath
drying
hot water baths
utting
crimping
washing
stretching
drying
cell wool
twisting
hardening, bleaching
rayon

Viscose process. The thick solution (viscose) is pumped through spinning nozzles into acid baths in order to form rayon filaments.

Photo USIS

When the bundles of filament extruded from a number of nozzles are grouped together into a "cable", cut up into lengths of only a few inches, and curled, the material known as cell wool (rayon staple) is obtained.

One of the products made from cell wool is motor tyre[1] cord, which is superior to cord made from cotton.

Rayon and cell wool produced by the process described above are still chemically the same substance that forms the main constituent of wood, namely the polysaccharide cellulose with the formula $(C_6H_{10}O_5)_n$.

Cuprammonium process

For the cuprammonium process and the acetate process it is necessary to use linters (short cotton fibres) or particularly pure wood cellulose (refined pulp). If pure cellulose, after undergoing a preliminary treatment with caustic soda and bleaching solutions, is treated with copper sulphate and ammonia, the macromolecular cellulose will, under the right conditions of concentration, dissolve as cellulose cuprammonium complex in aqueous alkaline solution. This viscous solution is filtered and extruded through spinnerets into a so-called spinning funnel. The filaments that are formed are drawn out in a powerful stream of water (wet spinning process), the cellulose is precipitated in the form of coherent filaments from the complex compound and is at the same time drawn out to very fine filaments—even finer than those produced by the silkworm. The washed product is dried and reeled up. After some further processing the final product is available as filament or cut staple.

Acetate process

Cellulose can be converted into numerous derivatives by chemical processing. These include cellulose acetate, which can suitably yield a serviceable synthetic fibre. Pure cellulose is acetylated with acetic acid and acetate anhydride in conjunction with sulphuric acid until the washed dry acetate becomes soluble in acetone. The solution is filtered and is extruded through spinnerets into warm air to form filaments. The acetone is thereby expelled from the filaments, carried away by the warm air, and recovered for re-use (dry spinning process). The dry filaments are reeled up, processed in various ways, and made available in the form of filament or cut staple.

1. Tire in U.S.A.

Fig. 1 THE CUPRAMMONIUM PROCESS

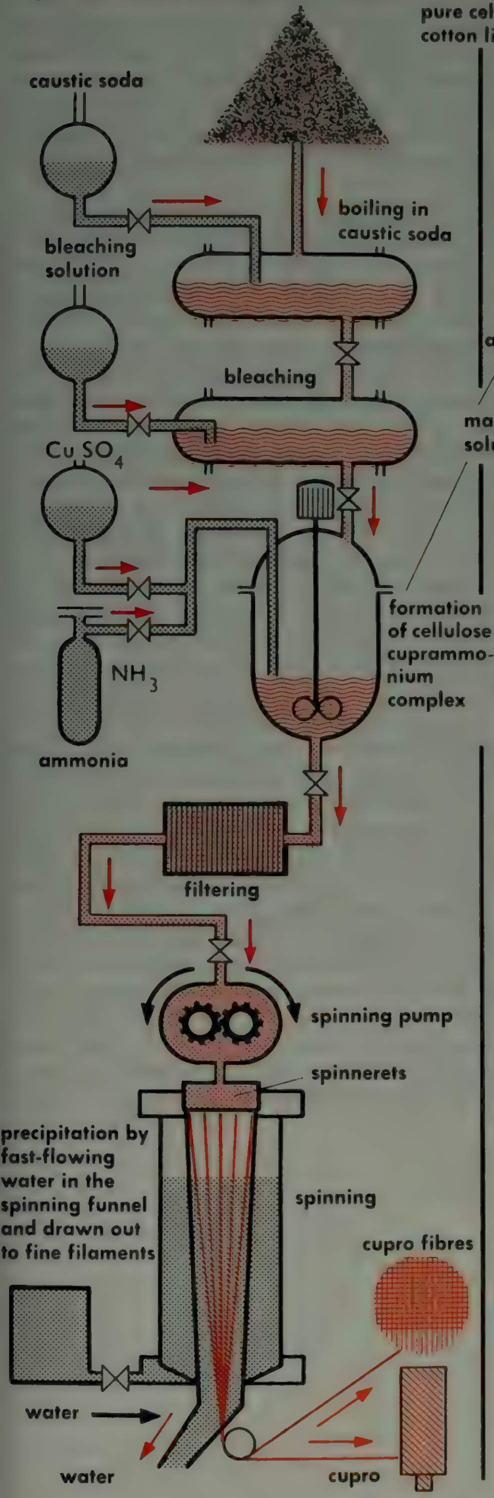
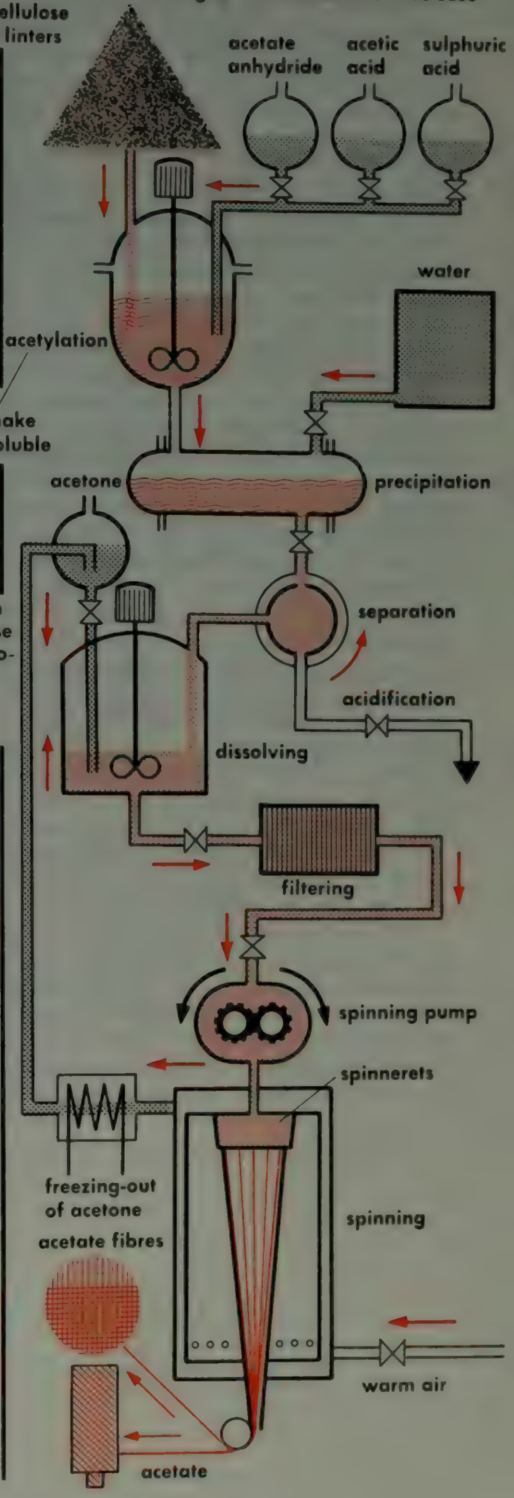


Fig 2 THE ACETATE PROCESS



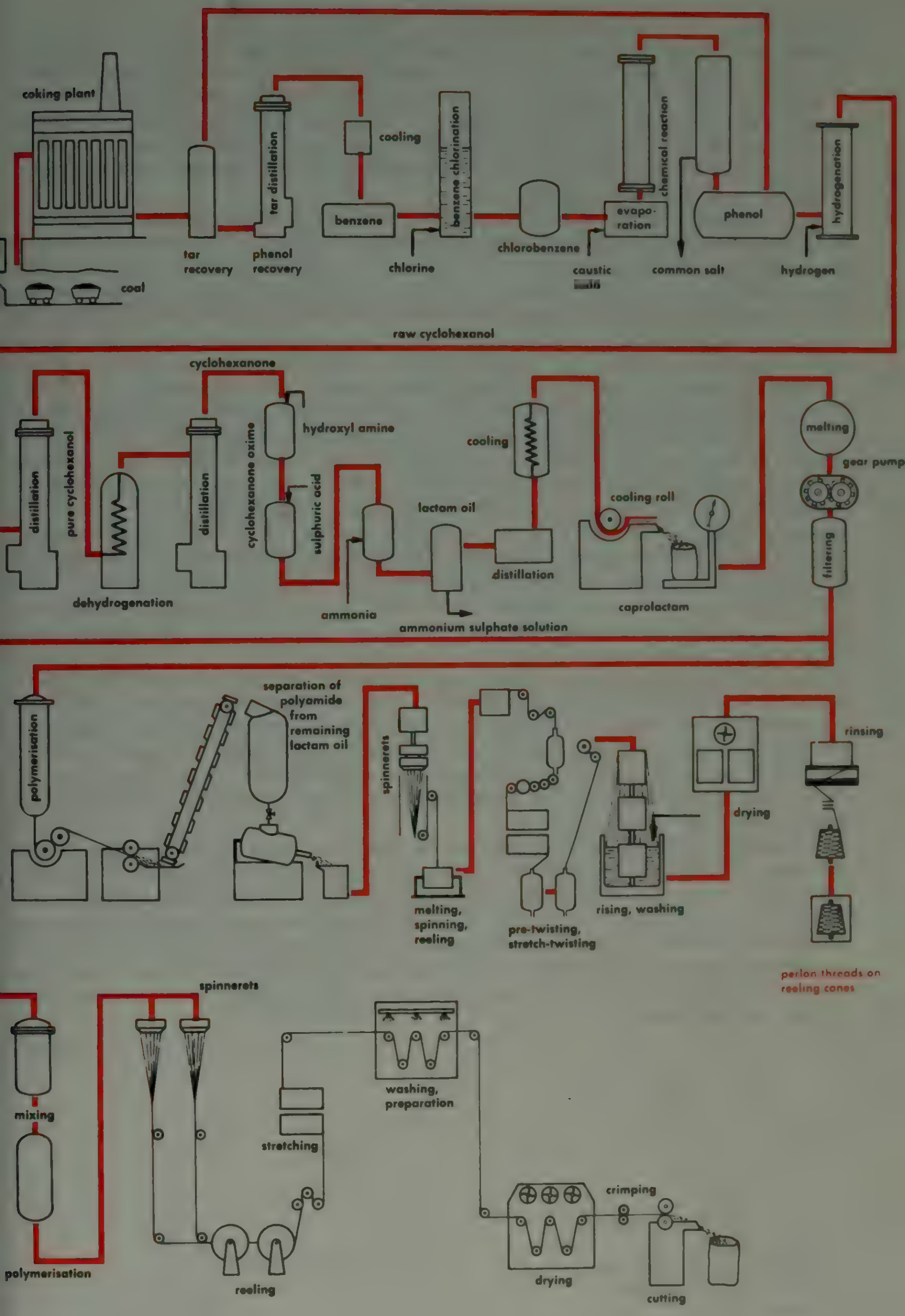
Cellulosic fibres are manufactured from raw materials such as wood and other woody vegetable matter, i.e., naturally formed materials whose molecules are merely rearranged by means of the viscose, cuprammonium or acetate process (see page 72 *et seq*). In the case of true synthetic fibres ("man-made" fibres in the true sense of the word), on the other hand, the molecules first have to be synthetically built up from the elements carbon, hydrogen, nitrogen and oxygen and formed into macromolecules. Four main groups of synthetic fibres are to be distinguished: polyvinyl, polyamide, polyacrylic, and polyester fibres.

The diagram shows diagrammatically, by way of example, the production process for the synthetic polymer called perlon, a German synthetic fibre resembling nylon. Both these products belong to the group of polyamide fibres which derive their name from the so-called amido group (CONH) in the molecule. Nylon and perlon are extensively manufactured from coal or petroleum, air, and hydrogen. Phenol is obtained from coal, and cyclohexanol from petroleum; these two raw materials are equally suitable for the purpose. Ammonia is produced from air and water, and oxygen is used in the manufacturing process. The substance eventually obtained, namely, caprolactam (for perlon) or the salt of adipic acid and hexamethylene diamine (e.g., for nylon 66), is polycondensed, with the result that the separate molecules become linked together to very long giant molecules. A substance which melts above 260° C is formed, which is filtered and is then extruded through spinnerets whereby the perlon (or nylon) filaments are formed. The filaments are drawn out, washed, dyed, and chemically processed to produce the finished synthetic fibres.

Perlon differs from nylon in that it contains up to 10% of monomeric constituents. For this reason the perlon melt remains stable, whereas the nylon melt is readily decomposable. From a given quantity of phenol as the raw material a larger quantity of perlon than nylon can be produced. Generally speaking, perlon and nylon have similar properties, but perlon has a greater affinity to dyes. Nylon, on the other hand, is better resistant to solvents. Fabrics made from these synthetic fibres are abrasion-resistant, tear-resistant and have better folding resistance than those produced from cotton or natural wool. They can withstand the action of putrefactive bacteria (for example, cotton fabrics buried in earth in which these bacteria abound will rot away completely in six months, whereas perlon or nylon will survive such conditions) and are also more resistant to attack by many chemicals than natural fibres are. Dirt does not penetrate into nylon and perlon fibres, and fabrics made from them therefore remain clean longer and are easier to clean than fabrics from natural fibres.

Both nylon and perlon are used for the manufacture of stockings, socks, sewing thread, upholstery fabrics, lingerie, carpets, artificial furs, bandages, surgical suturing material, nets and ropes for the fishing industry, motor-tyre[1] cord, straps, cable sheaths, filter cloths, clothing of all kinds, felt, brushes, sheeting, etc. In addition, they are used for a number of technical purposes, e.g., the manufacture of gear wheels and other components. Nylon was first obtained by the American chemist W. H. Carothers in 1932. Perlon was discovered by P. Schlack, a German, in 1938.

1. Tire in U.S.A.



coking plant
coal
tar recovery
tar distillation
phenol recovery
cooling
benzene
chlorine
benzene chlorination
chlorobenzene
caustic
evapo-ration
chemical reaction
common salt
phenol
hydrogenation
hydrogen
raw cyclohexanol
distillation
pure cyclohexanol
dehydrogenation
cyclohexanone
distillation
cyclohexanone oxime
hydroxyl amine
sulphuric acid
ammonia
lactam oil
cooling
distillation
ammonium sulphate solution
cooling roll
caprolactam
melting
gear pump
filtering
polymerisation
separation of polyamide from remaining lactam oil
spinnerets
melting, spinning, reeling
pre-twisting, stretch-twisting
rising, washing
drying
rinsing
perlon threads on reeling cones
mixing
polymerisation
spinnerets
reeling
stretching
washing, preparation
drying
crimping
cutting

Spinning is the process of making yarns or threads by the twisting of vegetable fibres, animal hairs, or man-made fibres, i.e., filament-like elements only a few inches in length. In the spinning mill the raw material is first disentangled and cleaned. Various grades and, perhaps, different kinds of fibre may be blended together at this stage so as to produce the correct quality of yarn. The fibres are spread out parallel to one ano her in a thin web, from which, in the next stage, a yarn-like material (called the "rove" or "roving") is formed. This can be done either by the web divider method or by the stretch-spinning process.

In the *web divider method* the raw material (fibres) is first untangled and straightened on a machine called a carding willow or breaker card (Fig. 1). Then comes the roving (preparatory spinning). The tufts of material are separated into individual fibres and the web is formed. The web is divided into narrow strips which are then rounded to produce the roving, this operation being performed by two or three carding engines (or cards) in a row. The material is fed at an accurately controlled rate to the first carding engine (Fig. 2) which further unravels the fibres and draws them accurately parallel by the action of the cylinder, worker rolls and clearers. In this way the so-called carded yarn is formed. On leaving the third carding engine the fibrous web is divided into strips by the divider (Fig. 3). These strips are then fed to the so-called rubbing gear, or condenser, which comprises a number of endless "rubbing leathers" which, in addition to a rotational movement, also perform axial to-and-fro movements, so that the web strips which are passed between them are rounded to form the roving. Next comes the fine spinning process. The rovings are drawn and thereby attenuated and more uniformly distributed. They are twisted into threads and are then wound on to reels. This is done either on a device called a self-actor mule (Fig. 4), or on a ring spinning frame (Fig. 5). The mule operates as follows: first the carriage travels out, the spindles revolve slowly, while the feed cylinder supplies the requisite amount of roving. The thin portions of the roving take up most of the twist and are thereby strengthened, while the thicker parts are attenuated. At each revolution of the spindle, the yarn springs off the end thereof and in this way acquires its twist. Then the carriage stops; the spindles revolve at high speed until the yarn has been twisted to the desired number of turns. Next, the spindles briefly revolve in reverse to uncoil the windings between the bobbin of yarn and the tip of the spindle. Finally, the carriage travels in again and winds up the spun yarn (intermittent spinning).

The ring spinning machine performs its task in a single operation (continuous spinning). The roving coming from the feed cylinder is pre-twisted by a drawing mechanism and is passed by the traveller to the bobbin which is fixed on to the spindle. When the latter rotates, the length of yarn between the spindle and the drawing mechanism is drawn taut, so that the traveller is pulled in the direction of rotation of the spindle. Because of its weight and the friction on the ring, the traveller rotates at a slower speed than the spindle. Each revolution of the traveller produces one turn of yarn on the bobbin.

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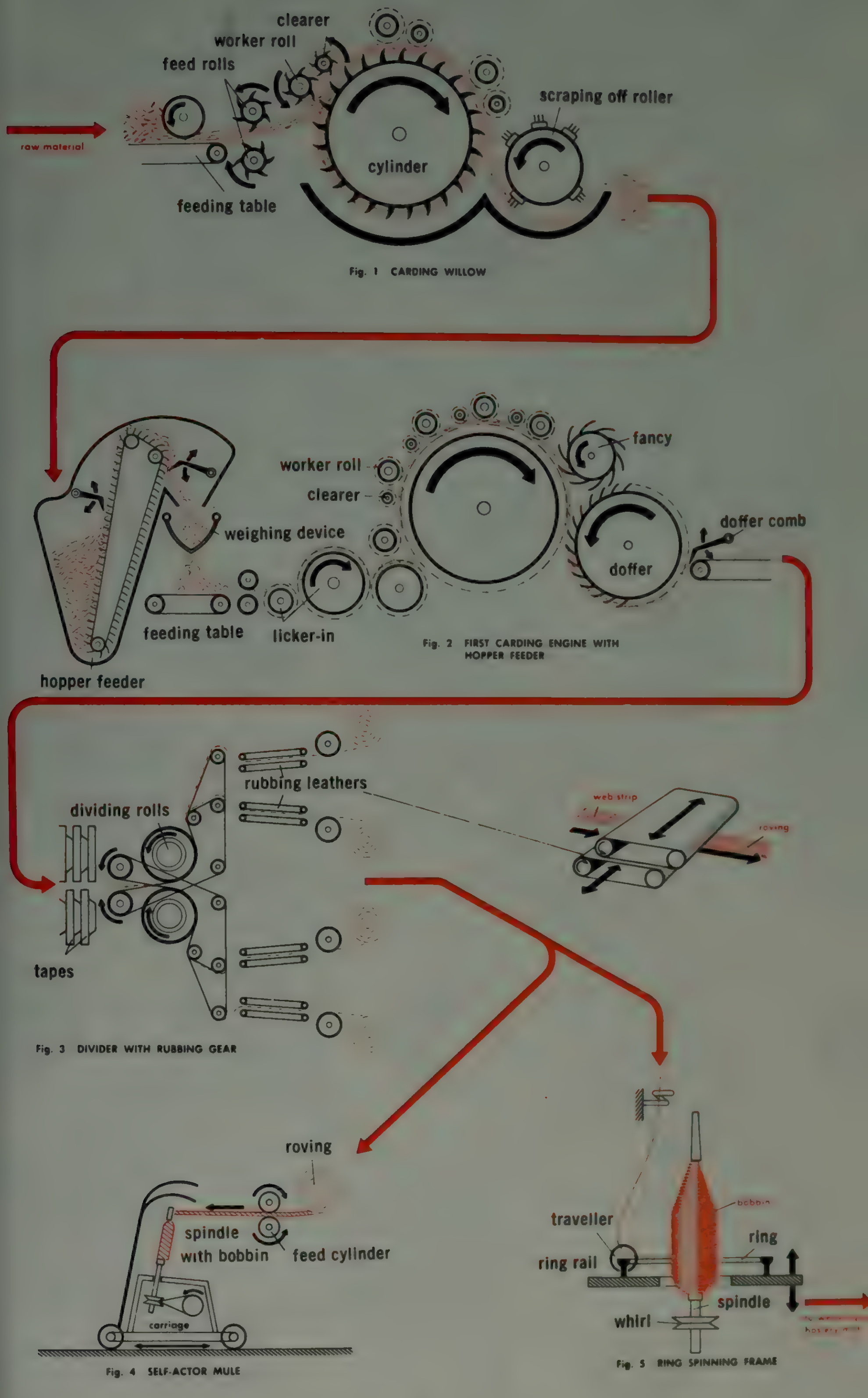


Fig. 1 CARDING WILLOW

Fig. 2 FIRST CARDING ENGINE WITH HOPPER FEEDER

Fig. 3 DIVIDER WITH RUBBING GEAR

Fig. 4 SELF-ACTOR MULE

Fig. 5 RING SPINNING FRAME

Part view of the spinning-mill at Derendingen, Switzerland
Photo Roland Schneider, Len Sirman Press

In the *stretch-spinning process* (here described more particularly with reference to cotton) the bales of wool are first loosened up on the hopper bale-breaker (Fig. 6) and then on the opener (Fig. 7). If necessary, different types of fibre are blended at this stage. The picker (or scutcher) forms the loose cotton into a sheet (the "lap") (Fig. 8). Separation of the fibre tufts into individual fibres and forming the web are done on a machine called a revolving flat card (Fig. 9), which is a particular type of carding engine. The taker-in roller (or licker-in) of this machine is set with teeth which tear away small bunches of fibre from the lap. These are stripped from the licker-in by a large cylinder with steel wire teeth. Over the cylinder are narrow bars (flats) carried by an endless belt and likewise provided with teeth which exercise a combing action and remove impurities. After leaving the card, the web is pulled through a funnel-shaped hole and is thus formed into a so-called "card sliver". To produce a yarn, the sliver has to be attenuated and twisted. Four, six or eight slivers are fed to the draw frame (Fig. 10); these are blended into one, and this operation is accompanied by attenuation, or drafting, so that the combined sliver becomes four, six or eight times as long but no thicker than the original card slivers. The draw frame comprises pairs of smooth rollers which are so driven that their speed of rotation increases from one pair to the next, until the front pair rotating at four, six or eight times the speed (depending on the number of slivers) as the back pair. It is this progressive increase in speed that produces the above-mentioned attenuation. To improve the regularity of the sliver, it is passed two or three times through the draw frame.

For producing very high-quality yarns, the card sliver is (before being fed to the draw frame) subjected to a process called combing in which up to 20% of the shorter fibres may be removed, thereby improving the spinning properties of the remainder. The yarn thus produced is called worsted yarn. The machine used is the comber (Fig. 11). The slivers are advanced so as to present a fringe to a set of revolving combs which remove the loose short fibres. The slivers are additionally drawn through a fixed comb. The combed fringe is then combined with the tail of the previous fringe.

The attenuation of the sliver after leaving the draw frame is completed on a series of drafting machines (known collectively as "speed frames") which operate on the same principle as the draw frame. In this way the roving is produced. At each stage, the drafted product must be twisted in order to increase its strength. This twisting is performed on flyer spindles (Fig. 12). The bobbin fits loosely on a vertical spindle to whose top is fixed the hollow inverted U-shaped flyer. The roving passes through the flyer to the bobbin, which it drags round. Because of the high speed of the spindle and flyer, the roving acquires a twist as it passes from the flyer to the bobbin. Fine spinning, i.e., the final operation in producing the finished yarn, is now usually done on a ring spinning frame or on a self-actor mule (however, the latter method is obsolescent for cotton spinning). These machines are similar in principle to those described with reference to the web divider method.

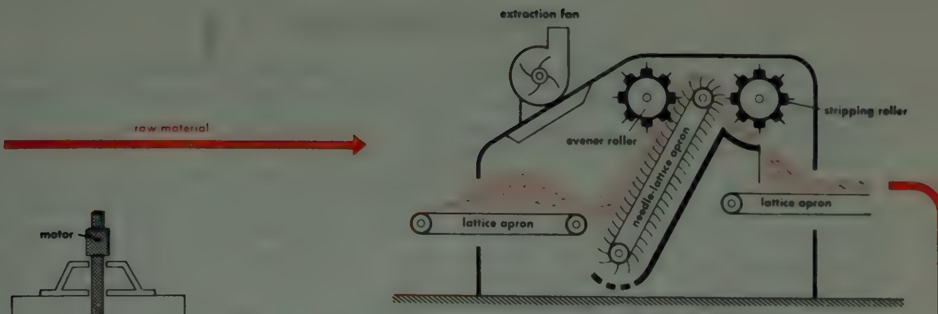


Fig. 6 HOPPER BALE-BREAKER

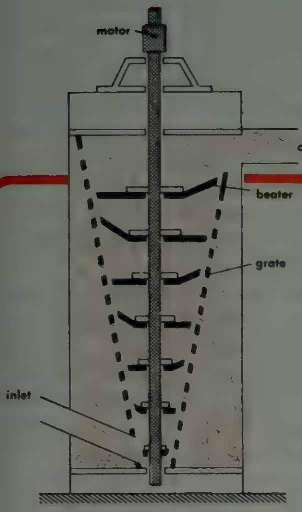


Fig. 7 OPENER

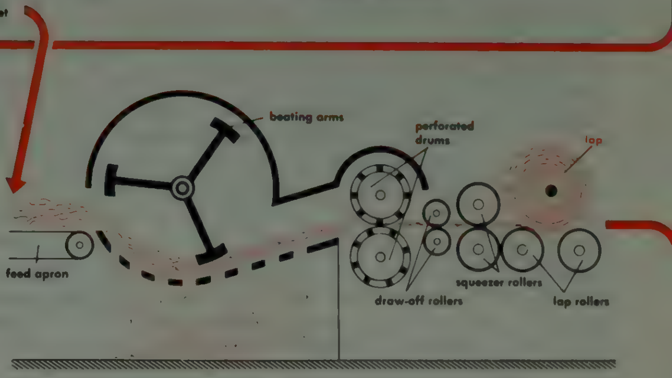


Fig. 8. PICKER

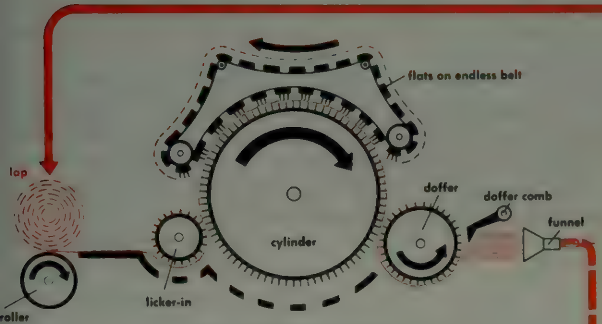


Fig. 9 REVOLVING FLAT CARD

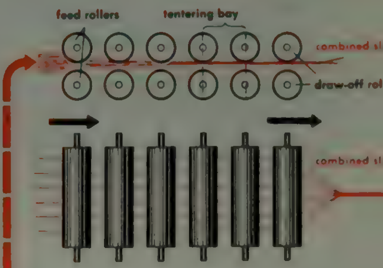


Fig. 10 DRAW FRAME

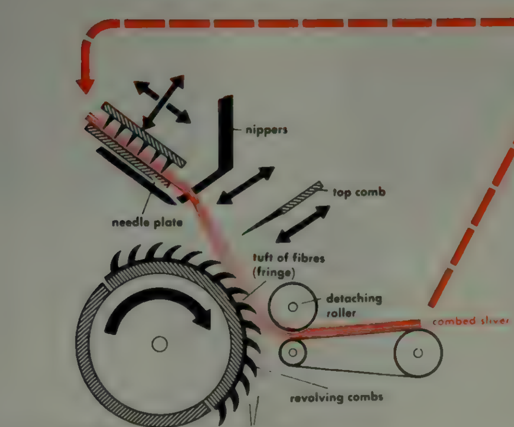


Fig. 11 COMBER

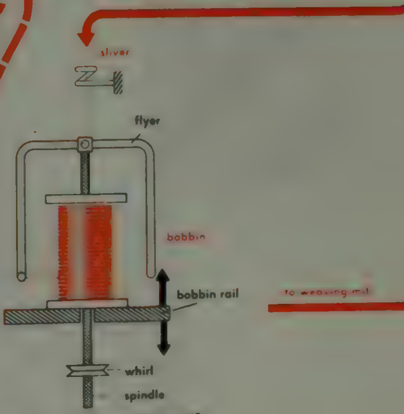


Fig. 12 FLYER SPINNING FRAME

For making textile fabrics (cloth) two systems of threads crossing each other at right angles are required. The longitudinal threads in a loom are called the warp, and the transverse threads are called the weft. The principle of weaving can best be explained with reference to the hand loom (Fig. 1). The warp threads, which are wound on the warp beam, pass over the back rest, through the eyes of the healds (or heddles), through the interstices in the reed behind which the shuttle plies to and fro, and then over the breast beam and on to the cloth beam (or cloth take-up roller). The shedding harness—comprising the heald shafts[1] with the steel wire healds and the treadles—forms the "sheds" for the shuttle to be passed through. The weaver works the treadles with his feet. When treadle 2 is depressed, shaft 2 (with the warp threads passing through the heald eyes) is lowered; at the same time shaft 1 (with the warp threads 1, 3, 5, 7, etc.) is raised. The shed is now open, and the shuttle containing the bobbin of weft thread is shot in. Next, the shed is closed, the swinging batten is moved forward, and the inserted weft thread is beaten into position in the cloth. Now treadle 1 is depressed, the shed is opened again; shaft 2 (with the warp threads 2, 4, 6, 8, etc.) is raised. The shuttle can again be inserted, and so on.

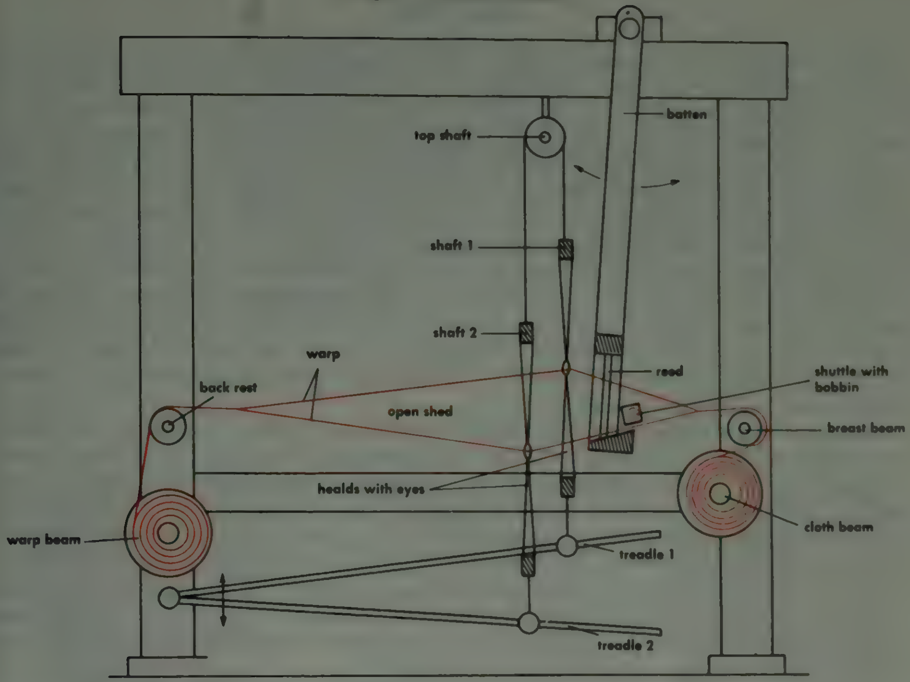
The warp and weft can intersect in many different ways. These can, however, be classified in three fundamental systems: (a) linen or plain weave; (b) twill weaves; (c) satin weaves. From these a large number of so-called "derived weaves" can be formed. The technical design of a weave is called its pattern. This is drawn on squared paper, on which the vertical lines of squares represent warp threads, whilst the horizontal lines represent weft threads. A filled-in square indicates that the warp thread it represents is above the weft, whereas a blank means weft above warp. Every weave has a certain number of "ups" and "downs" of the warp in relation to the weft. This can be represented by a symbolic notation. For instance, plain cloth is represented by 1/1 (i.e., one lifted, one depressed). Every pattern repeats itself. The area comprising the minimum number of warp and weft intersections constituting the pattern is called a "round of weave". In this area the warp threads are shown black; in the rest of the design they are in red.

The *linen weave* (L) has the smallest "repeat" of the pattern, L 1/1. It has a checkerboard pattern and presents the same appearance on both sides (see Fig. 2). The bottom diagram in Fig. 2 shows a section through the warp threads (small red circles) with the weft threads (in black) passing in and out between them. On the right, a similar section through the weft threads is shown. The linen weave is very widely applied (cambric, chiffon, cretonne, georgette, muslin, poplin, calico, etc.).

(Continued)

1. Heddles in U.S.A.

Fig. 1 HAND LOOM



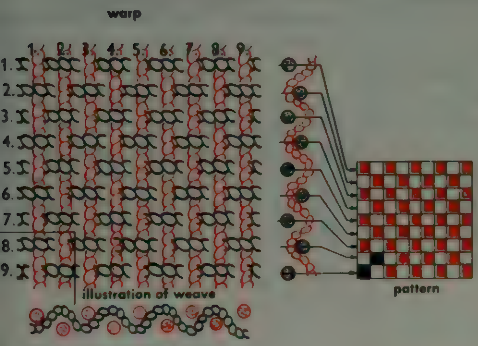


Fig. 2 Linen weave L 1/1

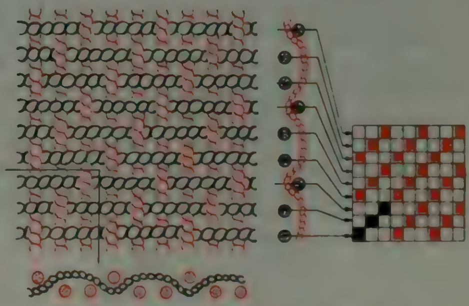


Fig. 3 Twill weave T 1/2 r

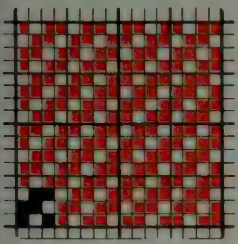


Fig. 4 T 2/1 r

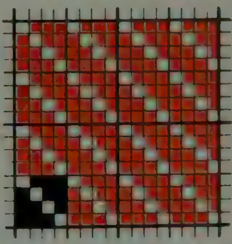


Fig. 5 T 4/1 l

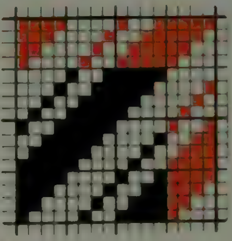


Fig. 6 T 5 1/2 4 r

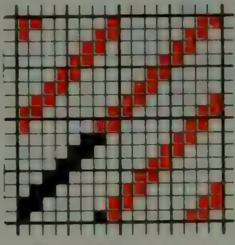


Fig. 7 T 2/5 r

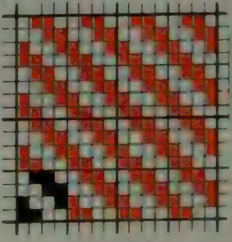


Fig. 8 T 2/2 l

Twill weaves (T) are characterised by diagonal ribs (twill lines). If these ribs run from bottom left to top right, the fabric is called right-hand twill (designated by r); if they run from top left to bottom right, it is called left-hand twill (designated by l). For equal density of warp and weft the ribs run at an angle of 45°. If the warp density is greater than the weft density, the ribs run at a steeper angle (steep twill), and vice versa (reclining twill). The simplest twills have the symbols T 1/2 r (Fig. 3) and T 2/1 r (Fig. 4). In the former there is more weft material on the face of the fabric (weft twill). Other twill weaves are shown in Figs. 5, 6 and 7. In the case of a double-faced twill (Fig. 8) the fabric presents the same appearance on both sides. A distinction is made between "simple" and "fancy" twills. In the former the same number of warp threads are placed successively above or below each weft thread, and the ribs are of uniform width (Fig. 8). In the latter more warp threads may be above one weft thread than another, the ribs may vary in width, and small ornaments may be introduced between the ribs (Fig. 6).

Satin weaves (S) differ from twills in having each warp thread lifted, or depressed, separately. Again a distinction can be made between warp satins and weft satins (or sateens). The simplest weft satin is S 1/4 (2) (Fig. 9), and the corresponding warp satin is S 4/1 (2) (Fig. 10). The distance from one point of intersection to the other is called the "progressive number" and is indicated by the figure in parentheses behind the symbol of the weave, e.g., (2), (3), (5). From five to upwards of thirty threads of warp and weft are required to complete the various schemes of intersecting. In satins and twills (as distinct from plain cloth) the finest or best threads can be made to predominate on the face ("right side") of the fabric. Some well known satin fabrics are charmelaine, duchesse, moleskin, velveteen, etc.

"*Derived weaves*" are obtained by modifying the fundamental weaves in various ways. Common variants of the linen weave are the *rib weaves* (repps) in which two or more warp or weft threads form ribs of various widths. In "weft repp" the ribs extend in the warp direction (Fig. 11) and the weft material is most in evidence on the face of the fabric; in "warp repp" the ribs extend in the weft direction (Fig. 12) and the warp material predominates on the face. Figured repps are formed by varying the width of the ribs (Fig. 13). Other popular variants of the linen weave are the *panama weaves*, in which two, three or four warp threads are lifted over, or depressed below, the same number of weft threads (Fig. 14). A figured panama weave is illustrated in Fig. 15.

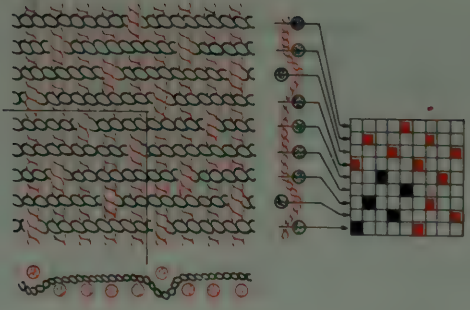


Fig. 9 Satin weave S 1/4 (2)

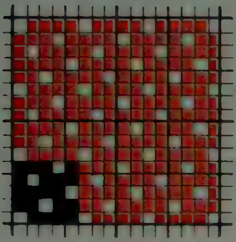


Fig. 10 S 4/1 (2)

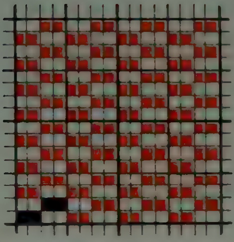


Fig. 11 Weft repp 1/1 2 f

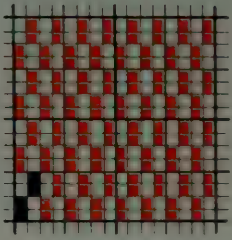


Fig. 12 Warp repp 2/2

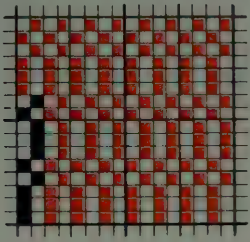


Fig. 13 Figured repp 3 1 1/1 3 1

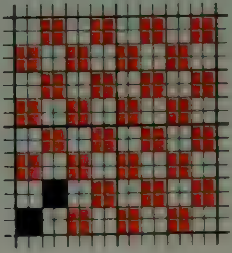


Fig. 14 P 2/2

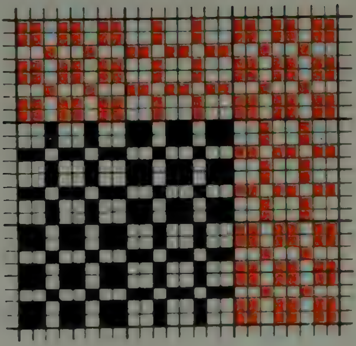


Fig. 15

Before weaving can commence, a variety of preparatory operations, involving a number of complex machines, have to be carried out. The most important of these operations are summarised in Fig. 1.

In the process known as *doubling*, two or more yarns are wound on to a bobbin without undergoing any twisting, as distinct from spinning, in which the fibres are twisted together to obtain the requisite strength (see page 80). In *twisting doubling*, however, two or more yarns (threads) are twisted round one another, the yarns being fed by means of delivery rollers, whereas in spinning the draft passes through a drawing frame (see page 84). Another preparatory treatment is *singeing* (see page 98).

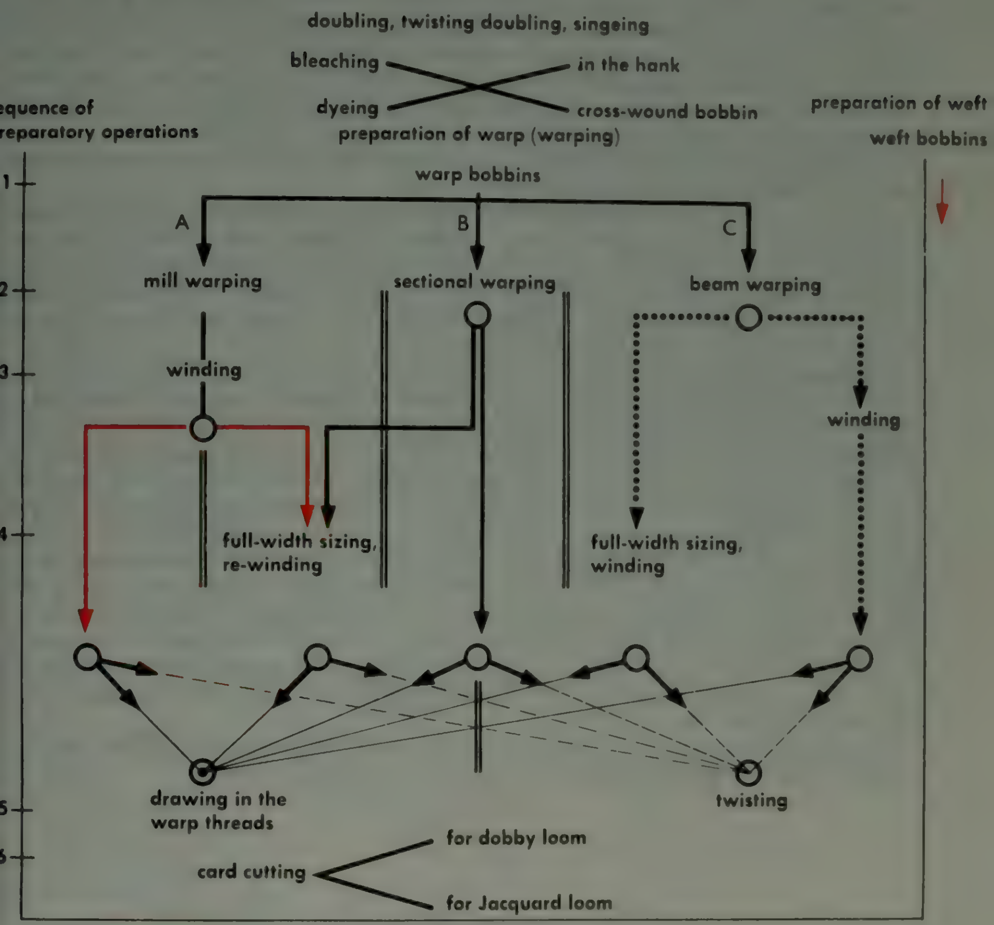
Bleaching and *dyeing* (see page 100) the yarns now tend to be done as the latest possible stage in the process because all the waste fibre material thus remains white and provides greater scope for utilisation.

Preparing the warp is called *warping*, i.e., providing a sufficient number of parallel threads. There are three methods: beam warping, mill warping, and sectional warping. A beam warper has a bobbin frame carrying some hundreds of bobbins. The threads are drawn separately through the interstices of an adjustable reed, then under and over a series of rollers, through the teeth of an adjustable comb, and on to a warp beam. The threads are thus drawn from the bobbins and wrapped in even coils upon the beam. In mill warping the warp threads are wrapped spirally round a very large reel rotating on a vertical axis. The distinctive feature of sectional warping (Fig. 2) consists in contracting the threads to form a 3–12 in. wide ribbon, which is coiled on to a warp cylinder. A number of these ribbons, or sections, are coiled side by side on this cylinder, which thus provides a means of intermediate storage. The threads are then transferred to a warp beam (loombeam), each section contributing its own width to that of the warp. Sectional warping is employed chiefly for coloured threads.

Sizing is a treatment for making the warp threads smooth so as to reduce the friction of the threads in the shedding harness (see page 86) of a loom. This is accomplished either by coating the threads or by saturating them with an adhesive paste (size). In one method the yarn for the warp is passed in the form of a sheet between a pair of rollers, the lower roller being partly immersed in liquid size.

Drawing-in, or entering, is the operation of passing warp threads through the eyes of a shedding harness in a sequence depending on the pattern to be produced. *Twisting* is the operation of joining the ends of a new warp with those of the previous one already in the loom. This is done by twisting the ends together, either by hand or with the aid of a special device.

Fig. 1 PREPARING WARP AND WEFT FOR WEAVING



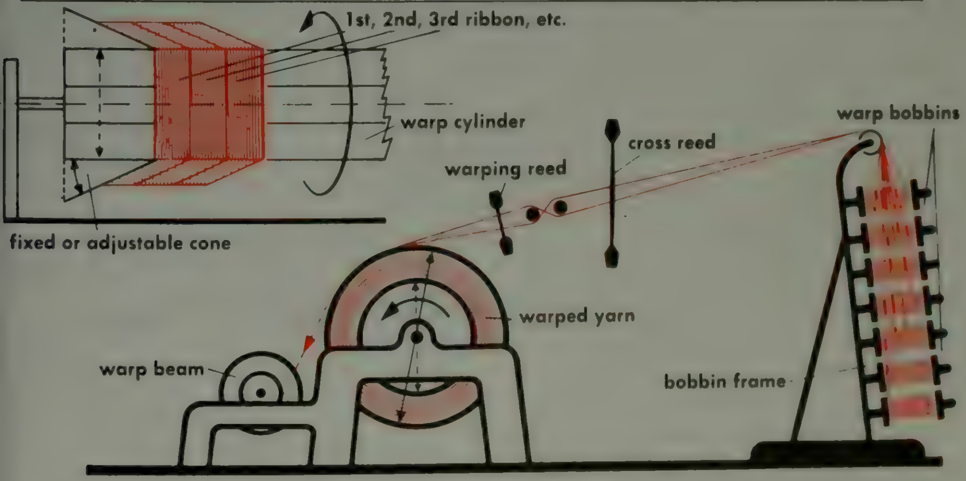


Fig. 2 SECTIONAL WARPER

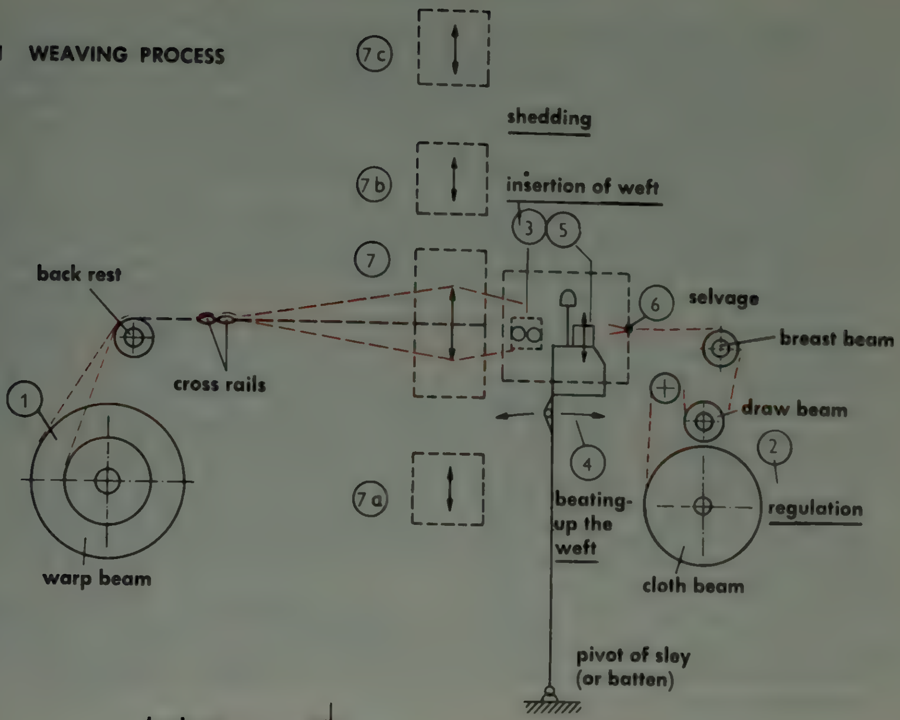
The power loom was developed in the late 18th and early 19th century. Its operating principle was, and still is, basically similar to that of the hand loom (see page 86). Hand looms have not been entirely superseded, however, and are still used for special purposes. The most important improvement applied to the hand loom was the development of the Jacquard machine in the 18th century. In a machine of this kind the warp threads are raised by rows of upright wires called hooks. The machine also comprises a series of horizontal needles which move to and fro. This system is used to operate the shedding and control the figuring of the fabric. Machines vary in size from 100 needles and hooks up to 1600.

Over the years a number of different loom types have been evolved for different kinds of work. As a rule, looms for the weaving of fabrics with small patterns are provided with healds for shedding (i.e., raising and lowering the warp threads in a predetermined sequence, so that they form a "shed" or passage for the shuttle; after a weft thread has been inserted, the shed is changed; cf. page 86); those for large patterns are provided with Jacquard equipment.

All these systems have certain main operations in common (a) shedding (see above); (b) inserting weft threads between the divided warp (this is done by the shuttle and is called "picking"); (c) beating-up, i.e., striking each weft thread into position in the fabric (this is done by the reed). In addition, the loom must have devices for holding the warp taut and delivering it as weaving proceeds, and for drawing away the cloth manufactured. In a power loom the above equipment acts automatically, and a number of additional features are provided to ensure trouble-free functioning. In Fig. 1 the main functional elements of a modern power loom are schematically indicated: (1) warp beam; (2) fabric take-off; (3) picking (by the shuttle carrying the weft bobbin); (4) beating-up of the weft; (5) shuttle-changing (with drop box or revolving box, so that yarns of different colours and/or qualities can be used); (6) forming the selvages on both sides of the cloth; (7) shedding: this can be done by means of tappets (7a), or by means of the so-called dobby mechanism (7b), or by means of the Jacquard machine (7c).

A post-war development is the jet loom (Fig. 2). This type of machine has no shuttle, the weft thread being carried along in a jet of air or water.

Fig. 1 WEAVING PROCESS



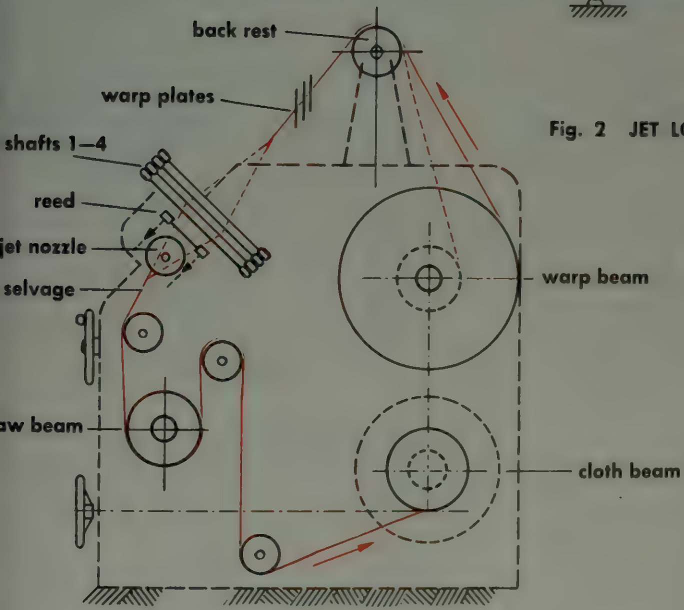


Fig. 2 JET LOOM (*side view*)

In British usage the term "hosiery" includes all types of knitted fabrics, whereas in the United States its meaning is restricted to stockings and socks. Knitting is used for the production of underwear and outer garments, curtain fabrics, lace fabrics, etc. The materials used are yarn and threads of cotton, wool and man-made fibres, also blended yarns and paper yarns. The products are either in the form of flat fabrics (subsequently made up into garments) or are ready-fashioned garments (pullovers, vests, stockings, etc.).

Hosiery is formed of looped fabric. There are two main types: weft fabric (Fig. 2) and warp fabric (Fig. 3). In weft fabric the threads extend crosswise across the fabric, whereas in warp fabric the threads extend lengthwise. The loops more particularly in a weft fabric form so-called "courses" (extending crosswise) and "wales" (extending lengthwise), as indicated in Fig. 2. For stockings and socks weft fabric is usually employed, as its elasticity or stretch is mainly crosswise and so provides a better fit for the leg. Warp fabric has less elasticity and is not much used for this purpose. The back of weft fabric is characterised by ridges running crosswise, the spaces between them being the courses. The face of the fabric displays the wales as a series of vertical lines. The back of the fabric is not so smooth as the face. On the other hand, warp fabrics are almost identical on the face and back. Some knitted fabrics derived from weft fabric are indicated in Figs. 4–6, and some derived from warp fabric in Figs. 7-9. An almost unlimited variety of design is possible, for example, by using combination warp and weft fabrics in which loops of the warp type are superimposed over weft fabric loops.

In the manufacture of full-fashioned women's hosiery the stocking blank is knitted in two pieces; the seam is a characteristic feature. Seamless stockings are produced on circular hosiery knitting machines. Other types of machine are also used for the purpose.

CIPLE OF LOOP FORMING

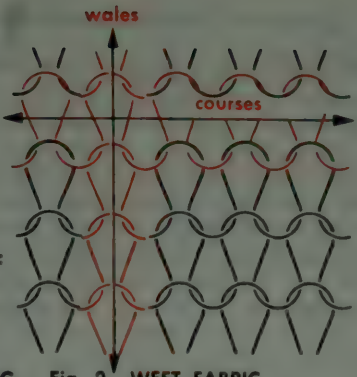


Fig. 2 WEFT FABRIC

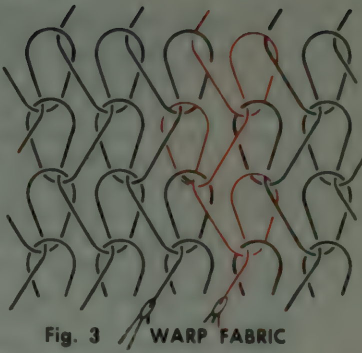


Fig. 3 WARP FABRIC

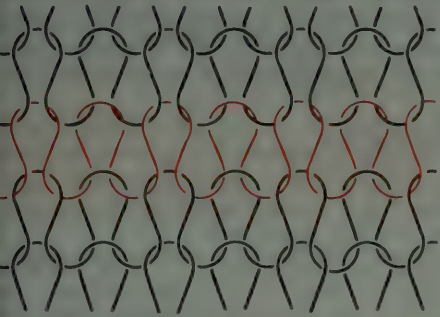


Fig. 4 RIB FABRIC

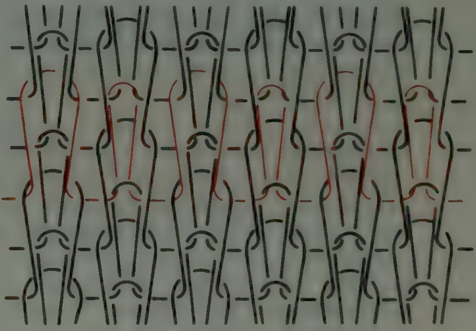


Fig. 5 INTERLOCK

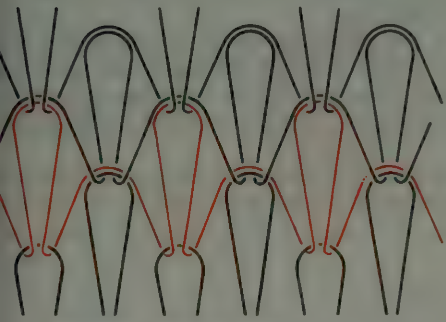


Fig. 6 TUCKED FABRIC

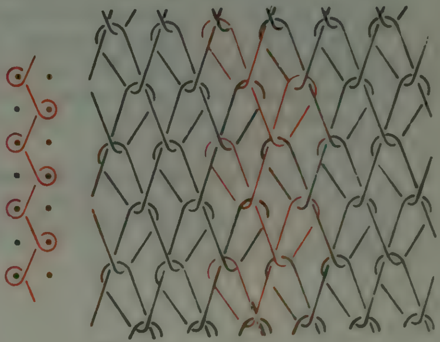


Fig. 7

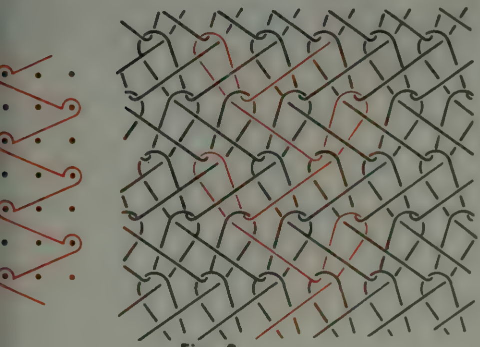


Fig. 8

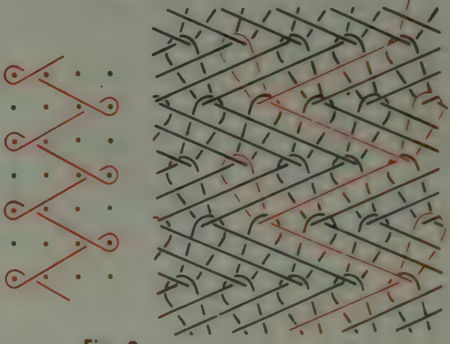


Fig. 9

The needles used for producing knitted fabrics on knitting machines are of two main types: the spring beard needle (Fig. 1) and the latch needle (Fig. 2). The operation of the beard needle is illustrated in Figs. 3a–d: The new thread, which is kinked around the shank below the open beard of the needle, slides past the open beard towards the hook, while the previously formed loop, resting a little below the tip of the beard, also slides upwards after the beard is closed over the new thread. The previously knitted loop thus passes over the beard and is moved towards the hook, so that the new thread in the hook is drawn through the previously formed loop, which is cast off as the new loop is formed. The latter then slides down the shank and past the open beard. The next loop can now be formed. Fig. 4 illustrates how the latch needle forms loops: The previously formed loop is on the shank of the needle, a little below the open latch. As the needle moves downwards with the new thread in its hook, the previously formed loop touches the open latch and swings it to the closed position. The previously formed loop can then be cast off over the closed hook, the new thread being formed into a loop by the hook. This new loop thereupon slides down, opens the latch, and finally slides off the lower end of the open latch on to the needle shank.

The best hosiery (women's stockings) is produced on the flat full-fashioned knitting machine (Cotton's machine, so named after the inventor). A modern machine of this kind may be as much as 60 ft. long and may comprise up to 32 sections, each of which produces a piece of flat hosiery fabric. Each section has about 400 needles set vertically. These needles perform a rather complex movement.

Hosiery produced on circular knitting machines, usually fitted with latch needles, is of the seamless type. The quality is coarser than that of stockings produced on full-fashioned flat machines. This is because latch needles cannot be constructed to such fine dimensions as spring beard needles. The stockings are shaped to fit the leg by varying the size of the loops, but the resulting fit of the tubular product is inferior to that obtained with full-fashioned hosiery.

Other types of knitting machine used for the manufacture of stockings are the circular rib machine and the Lamb type flat machine (Fig. 6). The latter can produce flat fashioned fabrics or it can produce seamless fashioned hosiery or other garments. The needles (usually of the latch type) are arranged in grooved beds placed at an acute angle to each other.

All the machines described can also be used for the manufacture of garments other than stockings. There are various other types of knitting machines used for particular purposes, e.g., the Milanese warp machine, the circular body machine, and the sinker wheel machine.

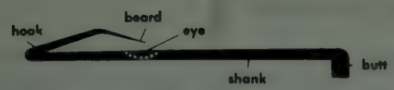


Fig. 1 BEARD NEEDLE

Fig. 2 LATCH NEEDLE

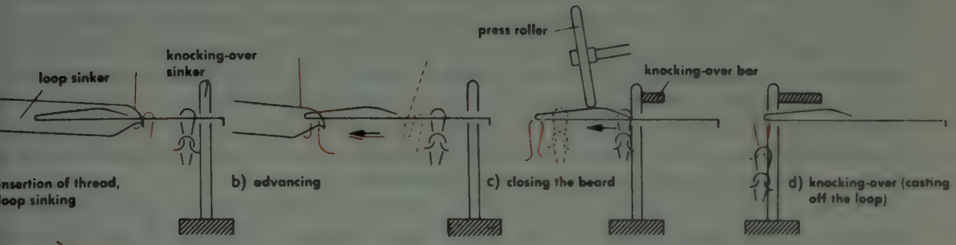
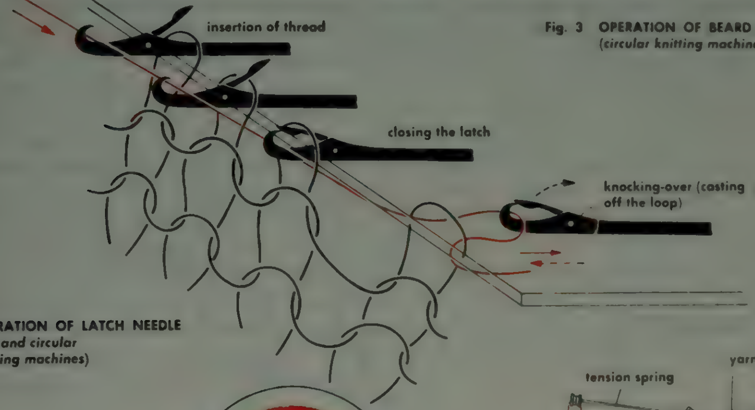
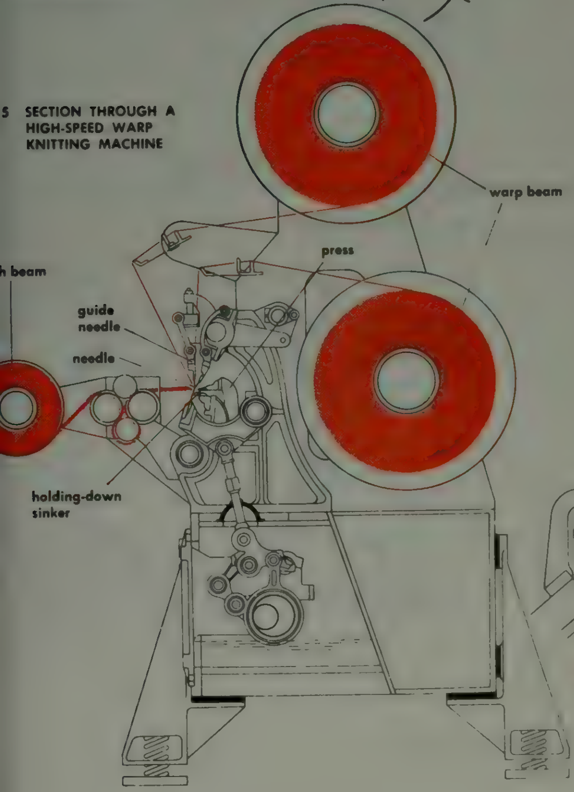


Fig. 3 OPERATION OF BEARD NEEDLES
(*circular knitting machine*)



4 OPERATION OF LATCH NEEDLE
(*flat and circular knitting machines*)



5 SECTION THROUGH A HIGH-SPEED WARP KNITTING MACHINE

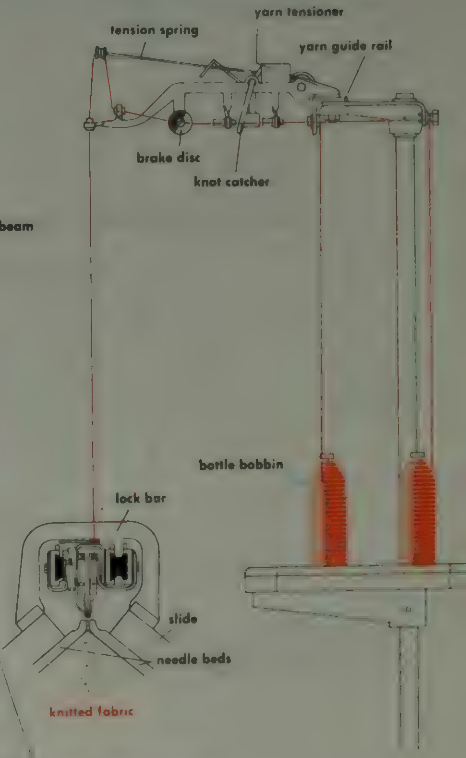


Fig. 6 LAMB TYPE FLAT KNITTING MACHINE

Most fabrics produced by weaving or knitting have to undergo a number of further processing treatments before they are ready for sale. In the finishing operations the fabric is subjected to mechanical and chemical treatment whereby its appearance and quality are improved and its commercial value thus enhanced. Each type of fabric has its own particular finishing operations. For example, textile produced from vegetable fibres require different treatment—raising, singeing, dyeing, printing, etc. —from those produced from animal fibres or synthetic fibres.

Woven cotton cloth is usually boiled with dilute caustic soda to remove natural oils and other impurities. Next, it is rinsed, scoured in an acid bath, processed in other ways, and then bleached, e.g., with sodium chlorite.

Singeing is carried out with the object of removing any fibres on cotton and rayon materials, more particularly if these have to be printed. The gas flame singeing machine (Fig 1) is usually employed for the purpose. In this machine a continuous web of fabric is moved past burners acting on the face or on both sides of the fabric. It then passes through a pair of squeezer rollers and a steam box which acts as a

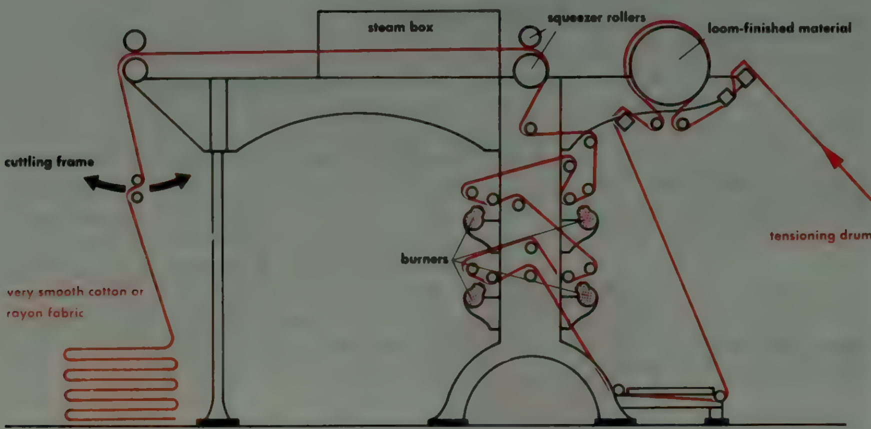


Fig. 1 FLAME SINGEING

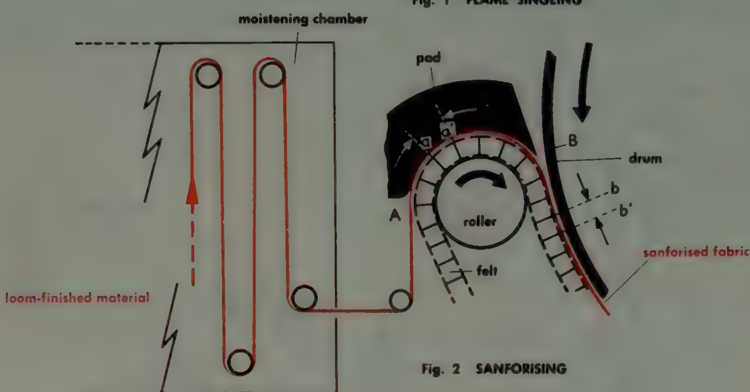


Fig. 2 SANFORISING

spark extinguisher. "*Sanforising*" is the proprietary name of a treatment which is applied more particularly to garments which are required to retain their shape well. In the course of spinning, weaving and finishing, the fabric is subjected to a good deal of pull and stretch, so that internal stresses are produced in it. When the material gets wet (rain or washing), these locked-up stresses are released and the fibres revert to their original length: the cloth shrinks. Sanforising produces a mechanical shortening of the fibres, so that it will not subsequently shrink in the wash. In this treatment (Fig. 2) the material first goes through the moistening chamber to soften the fabric. It then passes over a felt belt on to which it is firmly pressed by a heated pad. So long as the felt and the material are running over the roller together, they undergo a greater amount of stretch on the outside (between A and B) than on the inside. When the felt and material are then curved in the opposite direction (from the point B onwards), the stretched outer surface of the material is shortened (for instance a certain portion is first stretched to a length a–a′ and is then shortened to b–b′).

Raising (or napping) is a treatment in which small steel hooks (Fig. 3) tear some of the fibres, or the ends of fibres, out of the weft yarn, so that fabrics acquire a woolly surface called "nap". This improves their heat-retention and absorptive properties (as in flannel fabrics), besides making them softer to the touch. Certain types of raising treatment can improve the water-repellent properties of fabrics. Raising is used chiefly for cotton, rayon and woollen fabrics. It can be applied to one or to both sides. The raising machine, or napper, comprises a large drum provided with a number of peripherally mounted rollers which are covered with closely spaced steel hooks. All these rollers rotate in the same direction, but the hooks are curved in opposite directions on alternate rollers. The material passes continuously through the machine; the revolving brushes clean the napping rollers; the roller brush smoothes the material.

(Continued)

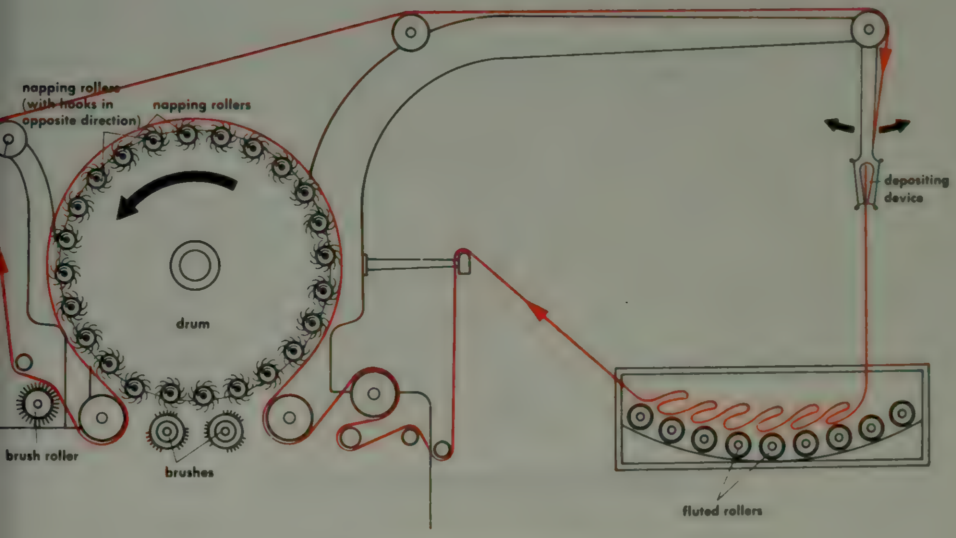


Fig. 3 RAISING MACHINE (24 *rollers*)

Textile printing: The dyes are dissolved in water. Thickening agents (e.g., starch) are added to the solutions in order to increase their viscosity. The oldest method is *block printing*, usually from wooden blocks in which the design is carved. *Stencil printing* is done with the aid of paper or thin metal stencils. Of greater practical importance is *silk screen printing*. The design is formed on a silk screen, usually by photographic methods, whereby a lacquer image is finally formed on the screen, which can then serve as a stencil. Most cloth is printed by *roller printing*, which is faster than other methods. The printing area is engraved on a copper roller so as to form a recessed pattern (intaglio), which is coated with the colour paste. Excess paste is scraped off by a steel blade. The roller then transfers the paste to the cloth. In multi-colour printing, a number of rollers, each printing a single colour, are arranged around the lower part of the central cylinder. Each impression must fall accurately upon the appropriate area within the complete pattern. Fig. 4 shows an eight-colour roller printing machine. The printing rollers may be hand-engraved, or etched, or the designs may be produced by photogravure methods.

The *dyeing* of textiles is a complex field of technology. There are two main types of dyeing machines. In one type the dye solution is circulated through the fabric, which is at rest; in the other, the fabric is passed through a stationary bath of the dye solution. Fig. 5 shows a machine of the latter type; the two squeeze rollers press the dye thoroughly into the fabric.

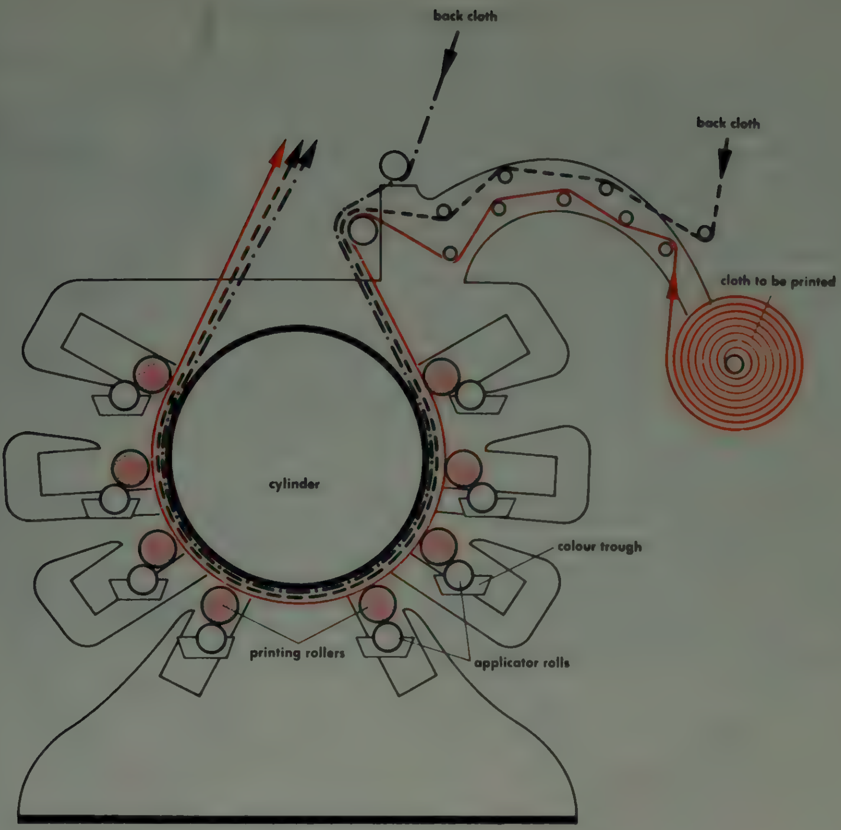


Fig. 4 ROLLER PRINTING MACHINE

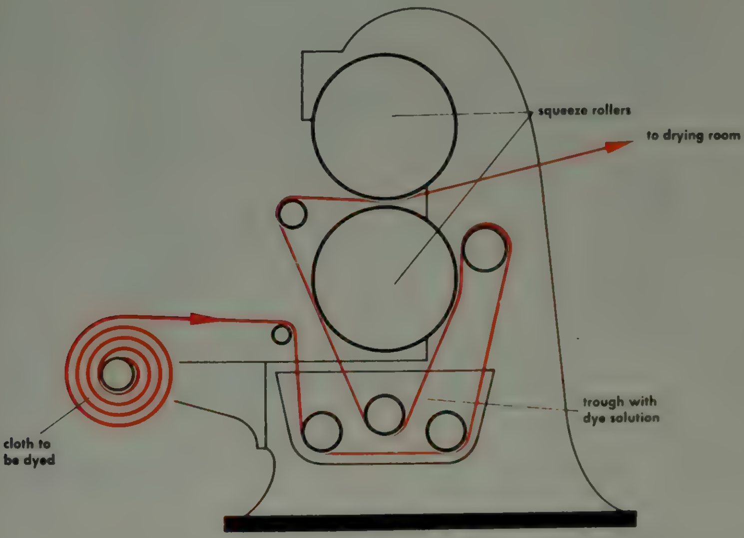


Fig. 5 DYEING MACHINE (*Foulard*)

ZIP FASTENER (ZIPPER)

The main parts of a zip fastener are the two chains of teeth, which are attached to strips of textile material, and the slide, which opens and closes the fastener.

Each chain consists of a large number of teeth, usually of metal, which are provided with small protrusions on the top surface and with corresponding recesses on the underside (Fig. 1). The protrusion on each tooth engages accurately with the recess in the tooth above. The two chains of teeth are slightly staggered in relation to each other. To close the fastener, the two chains must be so brought into engagement that the teeth on the two chains can interlock in pairs. This is achieved by the slide. At its upper end the slide comprises two divergent ducts, which join each other and merge into one duct as the lower end. The slide is so designed that the two chains of teeth are brought together at exactly the correct angle to make the protrusions interlock or one tooth engage with the recess on the opposite tooth (Fig. 2). At each end of the zip fastener are end pieces which prevent the slide from coming off. In some zips the two halves can be separated, in which case the bottom end piece is so designed that one chain of teeth can be withdrawn from the slide, while the latter is retained by the other chain.

Zip fasteners sometimes have plastic teeth (e.g., perlon), which are of a shape rather different from that of metal fastener teeth. The chains do not consist of individual teeth, but of loops formed by a spiral coil (Fig. 3). Fasteners of this kind have the advantage that, because of the resilient properties of plastic, they are not destroyed by tearing open. In addition to the types of zip fastener described above, there are many others, all of which operate on the same principle, however, and differ only in the particular form of the teeth employed.

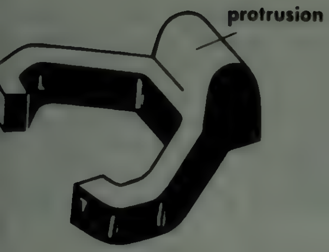


Fig. 1a

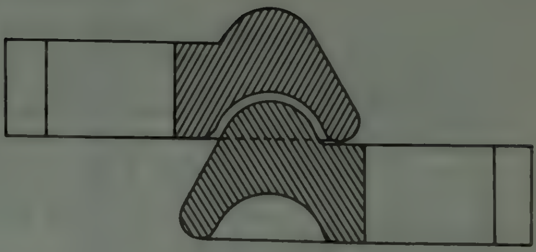


Fig. 1b

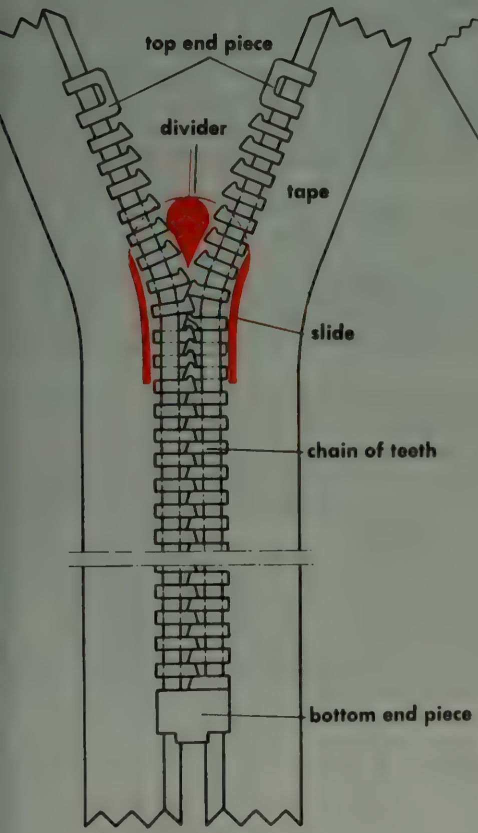


Fig. 2
ZIP FASTENER WITH METAL TEETH

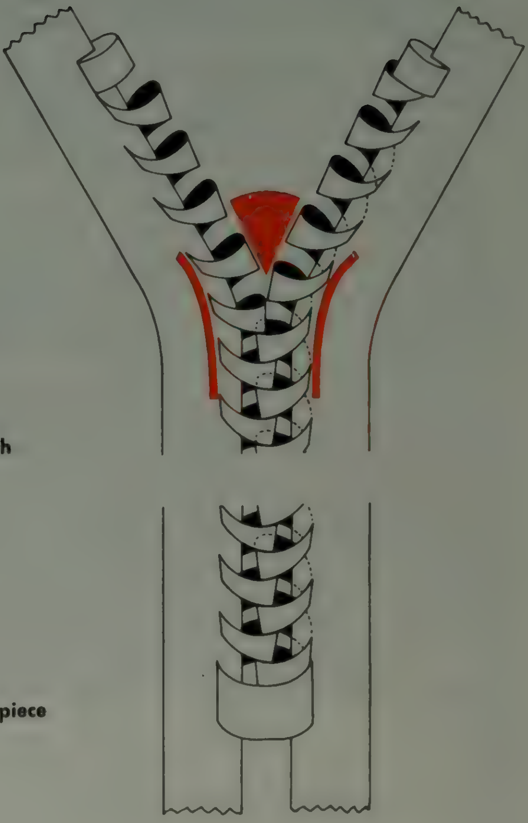


Fig. 3
ZIP FASTENER WITH PLASTIC TEETH

Conventional soap consists essentially of the sodium or potassium salts of higher fatty acids and has been known for hundreds of years. The drawback is that such soap washes effectively only in "soft" water (rainwater, water containing no calcium salts). In many districts the available water supply contains dissolved calcium salts ("hard" water), which form greyish deposits with soap, which cling firmly to the textile fabric.

It was particularly on account of the sensitivity of ordinary soap to hardness of water that the manufacture of so-called synthetic detergents was started over a hundred years ago. These are substances which develop an almost equally good washing action in hard and in soft water.

Synthetic detergents are the main ingredients of all washing agents and determine the washing power by their "surface-activity". This group comprises soap and other substances with soap-like behaviour which are for the most part synthetically manufactured from petroleum products or chemically decomposed fats.

Polyphosphates or complex phosphates are major constituents of any good washing agent. By forming complex compounds they neutralise the hardness of the water and combine with the heavy-metal salts, thereby increasing the washing power of the surface-active detergent. All washing preparations also contain bleaching agents, so-called "per-salts", which give off oxygen at temperatures above about 60° C. In the presence of suitable stabilising agents this oxygen bleaches any dirt or stain that is not removed by the detergent. This bleaching action is based on oxidation.

Whiteners are substances which convert ultraviolet rays into visible light and give the wash a brilliant whiteness—provided that it is really clean. Washing preparations also contain substances which increase the dirt-absorbing power of the lather. In addition, they contain fibre-protecting and dispersing agents, perfumes, colouring matter, skin-protective cosmetics, etc.

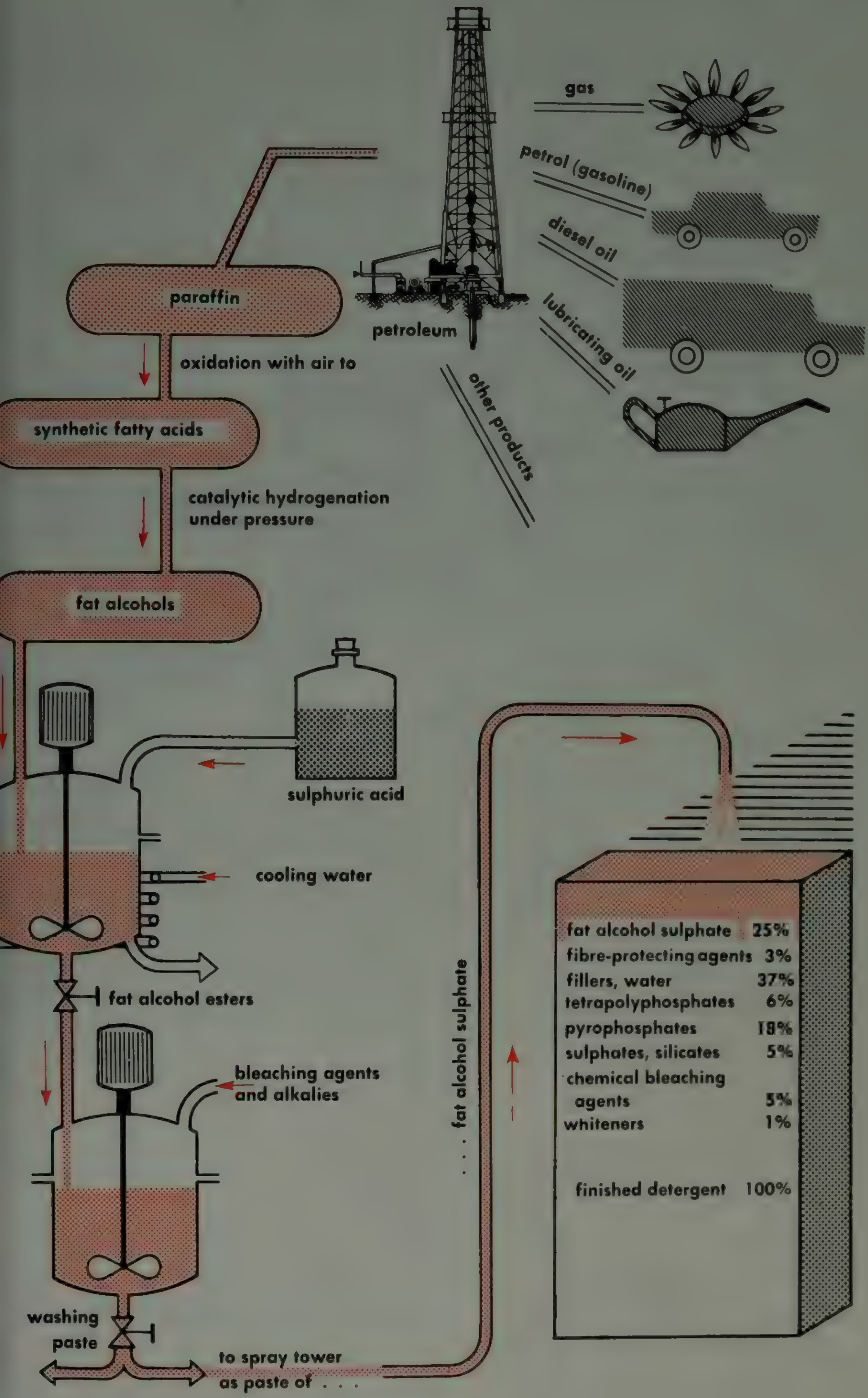
Besides washing agents intended more particularly for woollen or for boiling, there are various "special" preparations on the market, e.g., anion-active, cation-active and non-ionogenic detergents. These are intended for particular kinds of fibre or for dealing with particular kinds of dirt. Anion-active detergents are predominantly employed in household washing preparations. Non-ionogenic detergents are used to a less extent. Cation-active detergents have hitherto been used only for industrial purposes, e.g., in the textile industry.

Important anion-active detergents are alkyl sulphates, alkyl sulphonates, alkylaryl sulphonates. Alkyl phenol polyglycol ether and the polyglycol esters of fatty acids are examples of non-ionogenic detergents.

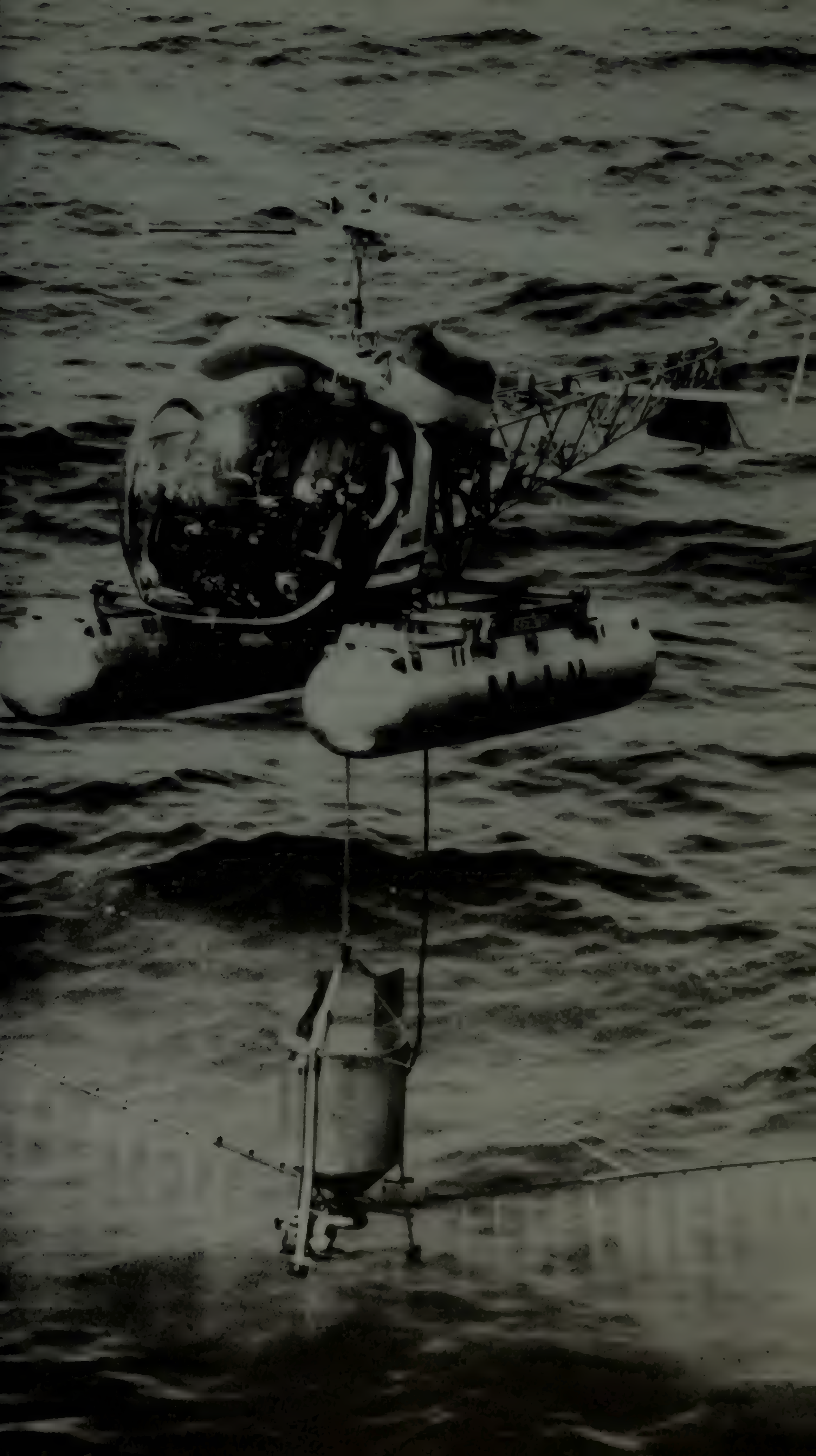
Wetting power, emulsifying capacity, dispersive and protective colloidal action, dirt absorbing capacity, and foaming power are the significant properties by which the performance of a washing preparation is judged.

Another requirement that washing agents will have to fulfil in the near future is that they should be biologically decomposable, i.e., they must be destroyed by bacterial action in watercourses and sewage purification plants, otherwise rivers and other waters will become increasingly polluted with foam, especially at weirs.

The various stages in the manufacture of a modern washing preparation are shown diagrammatically in the accompanying illustration.


gas
petrol (gasoline)
diesel oil
lubricating oil
other products
petroleum
paraffin
oxidation with air to
synthetic fatty acids
catalytic hydrogenation
under pressure
fat alcohols
sulphuric acid
cooling water
fat alcohol esters
bleaching agents
and alkalies
washing
paste
to spray tower
as paste of . . .
. . . fat alcohol sulphate
fat alcohol sulphate 25%
fibre-protecting agents 3%
fillers, water 37%
tetrapolyphosphates 6%
pyrophosphates 18%
sulphates, silicates 5%
chemical bleaching
agents 5%
whiteners 1%
finished detergent 100%

Detergent is sprayed by a U.S. petroleum firm's helicopter to "dissolve" oil spilled in the sea. Not harmful to marine life, the cleanser helps avert shoreline pollution.
Photo USIS



100 lb. of domestic washing is soiled with about 1.8 to 4 lb. of dirt as the result of normal use. This dirt consists mainly of fatty substances, proteins, dust and soot particles. Such dirt is very largely insoluble in water or is, indeed, water-repellent. Chemical analysis of this dirt shows that the following substances can (on an average) be extracted from 100 lb. of domestic washing: 0.9 lb. of protein-free organic matter (waxes, alcohols, hydrocarbons, etc.), 0.3 lb. of proteins (particles of skin, hairs, etc.), 0.15 lb. of fatty acids from sweat and greasy excretions, etc., as well as sand, dust and other inorganic constituents. A used shirt is soiled with greasy substances constituting up to about 0.25% of its weight; the collar may even contain as much as 1.2% of its weight of greasy dirt.

The dirt may be held by the fibres in various ways: mechanically (pigments are "jammed" between the fibres), chemically (fruit, oil or ink stains, etc., which in some cases can be removed only by bleaching), by absorption (feebler chemical bond, dissoluble by detergents), and by electrical forces.

As most of the dirt is relatively firmly held by the fibre, pure water by itself is not a very effective washing agent; it is necessary to add detergents (cf. page 104).

Very heavily soiled garments must first be left to soak in a detergent solution. However, this preliminary treatment is not usually necessary for normal domestic washing when modern washing agents are used. The actual washing process is determined by temperature, the mechanical treatment applied (rubbing, scrubbing, action of the washing machine, etc.), and chemical actions. The effect of these factors can be varied within certain limits; but on no account must any of these factors be overdone or neglected.

By adding a detergent to water, the surface tension of the latter is lowered: the water becomes more "fluid". The hydrophobic (water-repellent) end of the detergent molecule strives to escape from the water, while the hydrophilic (water-attracting) end wants to remain in the water. As a result, all the boundary surfaces of the water (e.g., between the water and the air or the wall of the wash-tub or the clothing itself) are densely packed with molecules of detergent. In their effort to escape from the water, detergent molecules penetrate in between the dirt and the fibre, where the water alone could not go. Aided by the mechanical actions applied (scrubbing, etc.), the detergent molecules manage to push farther and farther between the dirt and the fibre until in the end the dirt is completely dislodged. By the same process, particles of dirt which are stuck together, as well as oil and grease, are finely divided and are absorbed by the lather, which must retain the dirt in suspension and not re-deposit it on the clothing before the latter is finally rinsed with clean water.

Fig. 1 DETERGENT DISSOLVES IN WATER TO GIVE NEUTRAL SOLUTION

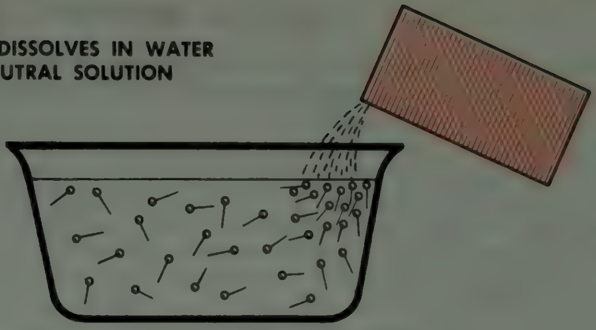


Fig. 2 DIAGRAM OF DETERGENT MOLECULE

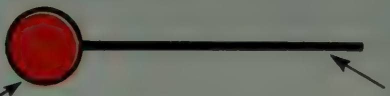


Fig. 3 DETERGENT ACTS UPON DIRT PARTICLES ON THE TEXTILE FIBRE

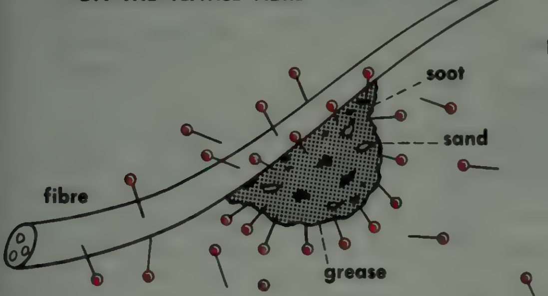


Fig. 4 DIRT PARTICLES ARE DISLODGED

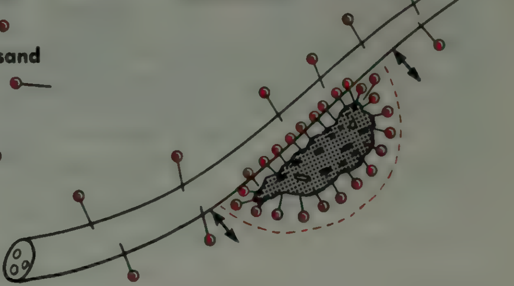


Fig. 5 DIRT PARTICLES ARE COMPLETELY ENVELOPED (*so that they become "soluble" in water*)

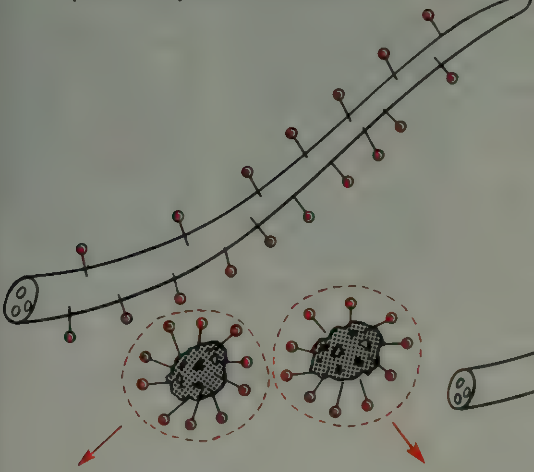
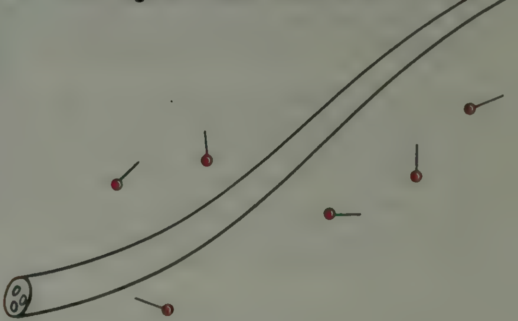


Fig. 6 CLEAN FIBRE IS RINSED



There are two main types of washing machines: (a) Tub type machines (e.g., agitator machines, nozzle machines, paddle-wheel machines). (b) Drum type machines (fully automatic, semi-automatic).

In all tub type washing machines the load to be washed is moved about suspended in water. In an agitator machine the movement is produced by fins or blades slowly revolving on a central shaft (Fig. 1). In a nozzle machine it is produced by water jets streaming from nozzles at the bottom of the tub, and in a paddle-wheel machine it is produced by a flat paddle wheel rotating at high speed. Tub type machines are provided with a draining pump, are often independently heatable by electricity, and usually embody semi-automatic features, i.e., the temperature and the washing time can be pre-set to required values. When the desired temperature is reached, the agitator mechanism or paddle wheel is started automatically and continues to run for the predetermined time. Water is extracted in a separate power-driven spinner or wringer.

In the drum (or cylinder) type machines (Figs. 2a and 2b) a perforated steel drum is rotated inside a tank. In some machines the drum turns in one direction only,

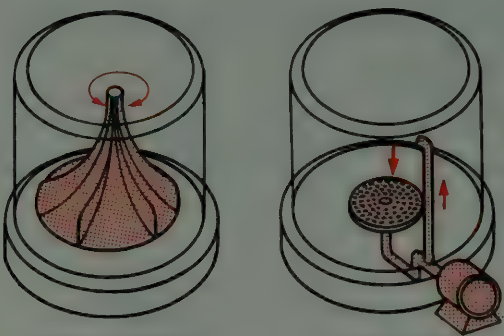


Fig. 1 TUB TYPE WASHING MACHINES

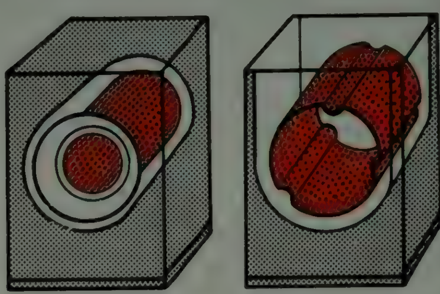
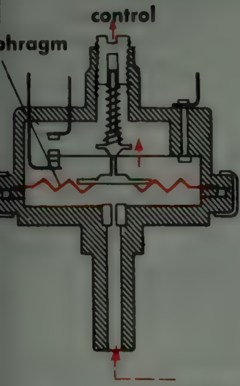


Fig. 2 DRUM TYPE WASHING MACHINES



3 **PRESSURE-OPERATED CONTROL**

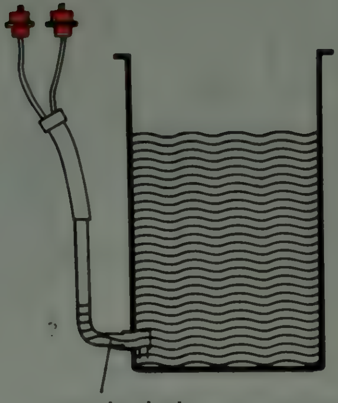


Fig. 4

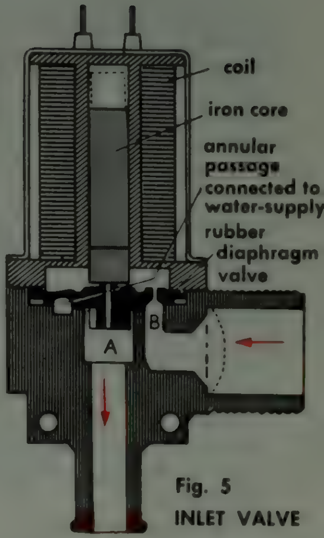


Fig. 5 INLET VALVE

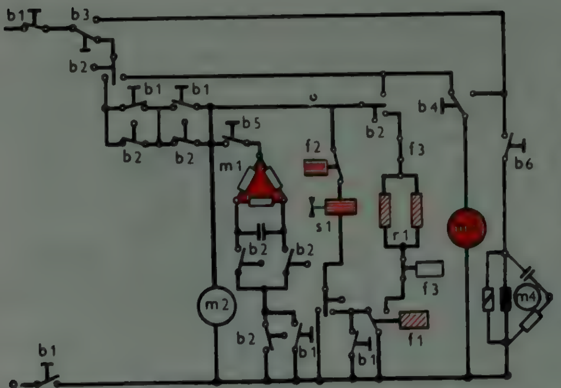
whereas in others it reverses at intervals. Projections (baffles) inside the drum cause the clothes to drop back into the water as the drum revolves. A fully automatic machine of this kind pre-soaks, washes, rinses and extracts water with no attention from the operator after the load is placed in the machine, the power switched on and the timing device set. A semi-automatic machine has to be filled by hand (through a hose connected to the water supply), and at the end of the washing operation the water must be drained from the tank by switching on a pump.

The following is a description of the principal components and the mode of functioning of a fully-automatic washing machine.

The inflow of water and the water level in the tank is controlled by a pressure-operated control device (Fig. 3). It comprises a diaphragm which is forced upwards as the pressure under it increases and which actuates a contact when the predetermined pressure (and therefore depth of water) has been reached. The contact completes an electric circuit which causes the water inlet valve to close. Different quantities of water are required for rinsing and for washing, and for this reason there are usually two control devices. These are connected to a side tube communicating with the tank. As the water rises in the tank, it also rises in this tube and compresses the air in the upper part thereof. It is this air pressure that actuates the diaphragms in the control devices (Fig. 4).

The inlet valve (Fig. 5) is controlled by the control device. When no current is flowing, the iron core rests on the rubber diaphragm and keeps the opening A closed. As the active area upon which the water pressure acts above the valve is larger than below it, the diaphragm is pressed firmly on the valve seat, so that the water, by its own pressure, keeps the valve closed. When the coil is energised, the magnetic core is pulled into the coil, so that A opens. Equalisation of pressure now occurs through the openings A and B. The pressure acting on the rubber diaphragm from below increases and lifts the diaphragm off its seat.

The programme control unit emits the electric impulses which control the various operations that the machine performs (Fig. 6). It comprises the drive, reversing mechanism, washing time selector, step-by-step switching device, and programmed switching device. The control unit as a whole is driven by a synchronous electric motor which drives various cams, each of which has a particular shape that controls



b1—selector switch
b2—programmed switching device
b3—change-over switch for washing/spinning
b4—change-over switch for draining pump
b5—switch for locking the cover (washing motor)
b6—switch for locking the cover (spin-dryer motor)
f1—pressure-operated control device (low level)
f2—pressure-operated control device (high level)
f3—thermostat
r1—heating system
s1—inlet valve
m1—washing motor
m2—synchronous motor in programmed switching device
m3—draining pump
m4—spin-dryer motor with magnetic brake

Fig. 6 ELECTRICAL CIRCUIT OF AN AUTOMATIC WASHING MACHINE

a certain sequence of operations. The edges of the cams are "scanned" by spring contacts and thus, for example, cause the drive motor of the washing drum to reverse its direction of rotation at intervals (Fig. 8) or determine the length of the washing and/or rinsing time, which can be varied as desired by means of the washing time selector.

The purpose of the step-by-step switching device is to transform the rotation of the synchronous motor into a stepwise motion for the programming and control of the washing and rinsing operations. A control diagram of the programme control unit is shown in Fig. 9: the heating system and the inlet valve (low level) are not switched on during the whole of the time span indicated. The heater is switched on and off by a thermostat (p. 26, vol. I), while the inlet valve is actuated by the pressure-operated control device. The washing temperature can be pre-set to any value between 30° and 100° C, and the thermostat keeps the actual temperature to within a degree or two of this selected value. The waste washing water and the water extracted from the clothes by spin-drying is removed by means of a draining pump. For water extraction by centrifugal action (spin-drying) the drum is rotated at high speed (about 2800 r.p.m.).

Modern washing machines are equipped with a number of safety devices. The rotation of the drum is automatically switched off when the cover is opened, and a magnetic brake quickly stops the motion, so that there is no danger of the operator's hand being caught in the revolving drum. Another protective device prevents the tank being heated unless it contains water up to the correct level. In the event of an interruption of the water supply to the machine during operation, a safety device stops the programme and switches off the heating.

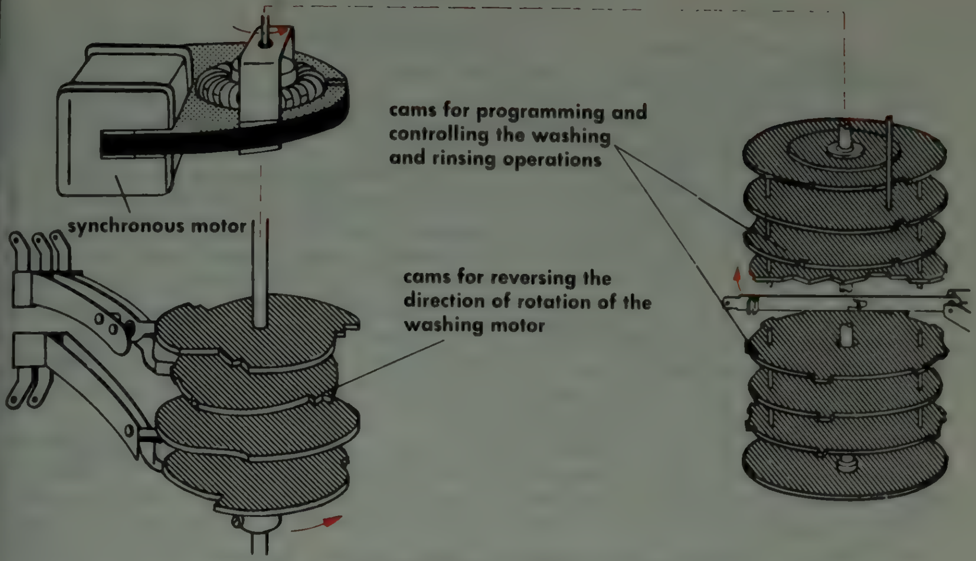


Fig. 7 PROGRAMMED SWITCHING MECHANISM

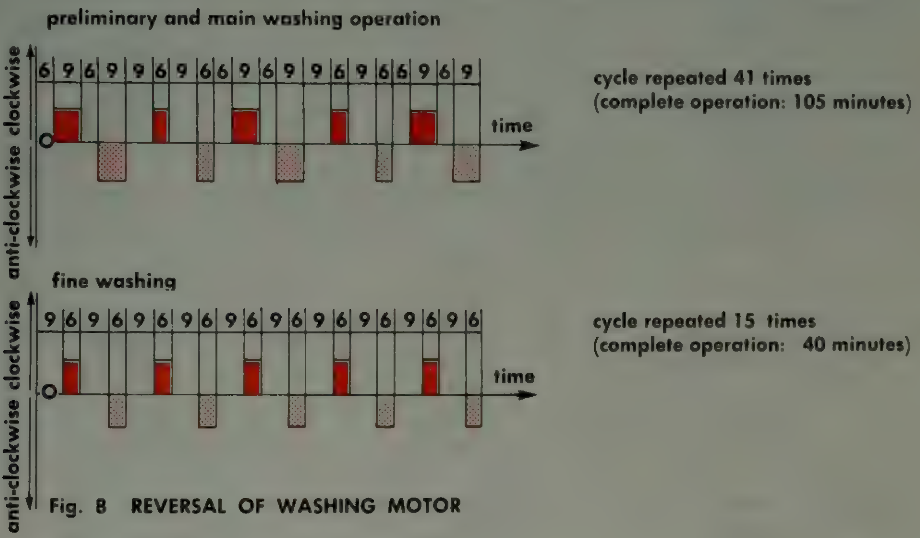


Fig. 8 REVERSAL OF WASHING MOTOR

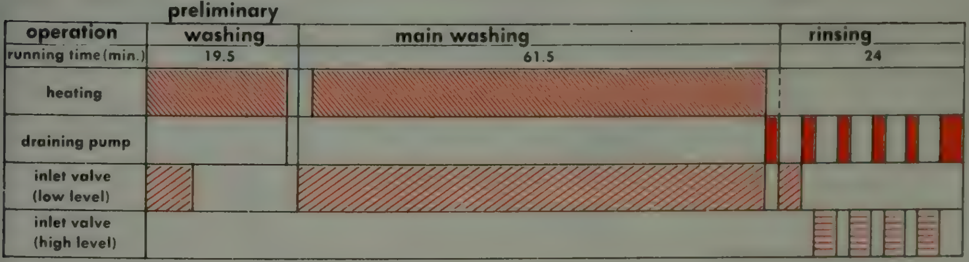
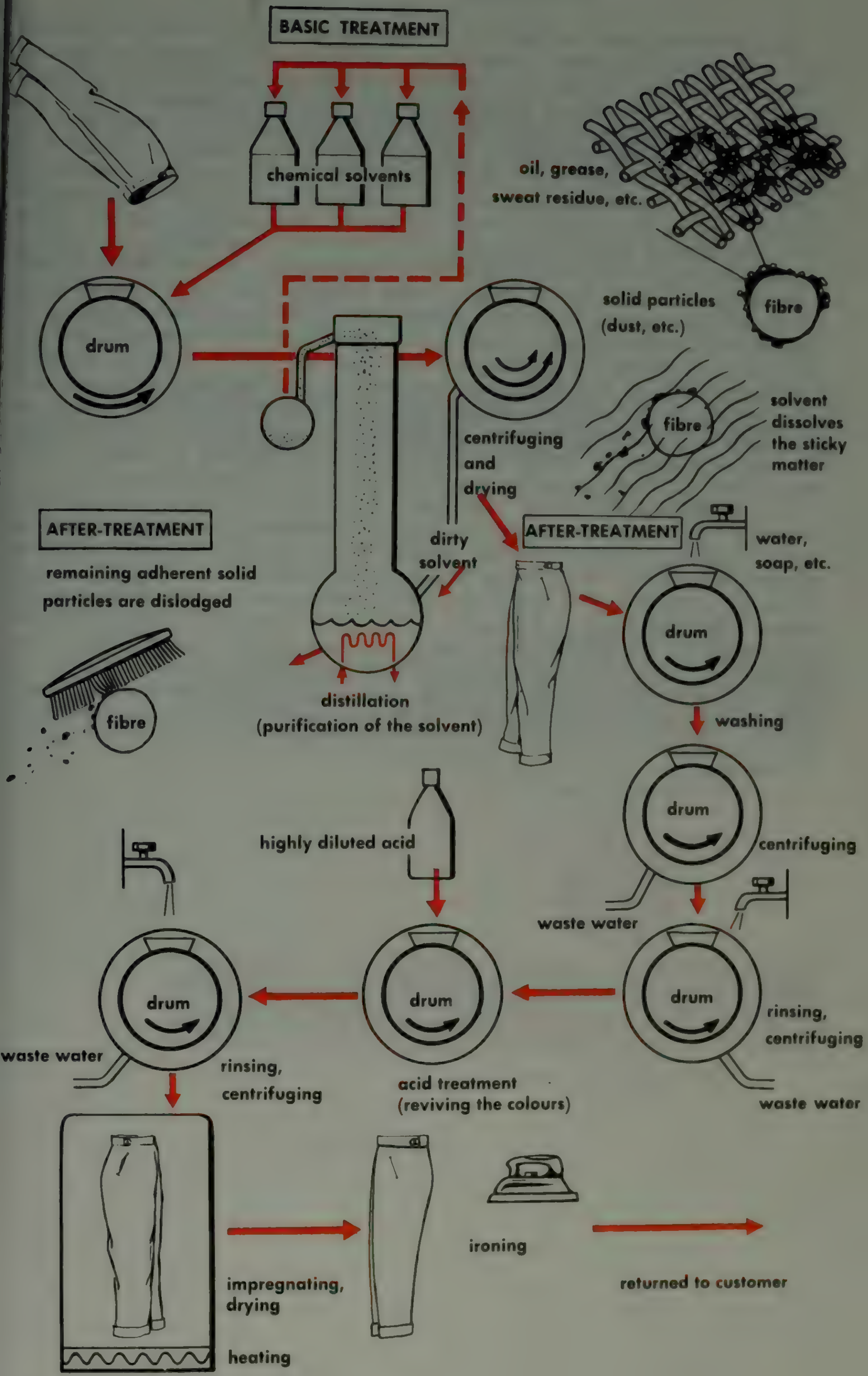


Fig. 9 CONTROL DIAGRAM OF PROGRAMME CONTROL UNIT

In most cases dirt remains adhering to the fibres of textile fabrics by two kinds of forces: first, by sticking to a coating of grease or a dried coating of substances which swell up in water or other solvents (e.g., starch, proteins and other glue-like substances); second, by direct adhesion because of the physico-chemical character of the fibres and the dirt (cf. page 108).

"Dry" cleaning (more properly called "chemical cleaning") uses liquids other than water for the cleaning of fabrics. In this treatment, adhering dirt of the first kind is removed by dissolving the grease or other sticky matter to which the dirt particles are clinging. A wide range of solvents are employed: carbon tetrachloride, trichloro-ethylene, tetrachloro-ethylene, naptha (petroleum ether), benzene, etc. (Carbon tetrachloride is a toxic substance and its use has largely been discontinued). A modern dry cleaning plant comprises a number of specialised machines and appliances. The soiled garments are treated with the solvents in rotating drums. The contaminated solvent is drained off and purified for re-use. The cleaned garments are dried, impregnated (if desired), and re-shaped.

In some cases, however, it is necessary to use water as an additional solvent or swelling agent. The second type of dirt is dislodged from the fabric by means of detergents added to the water. After this "wet" treatment the fabric is usually treated in very weak acid solutions (to revive the colours), rinsed, centrifuged and dried. Impregnation treatment may be applied at an intermediate stage to stiffen the fabric and make it water- and dirt-repellent. Garments are finally pressed on special machines operated with steam and air.

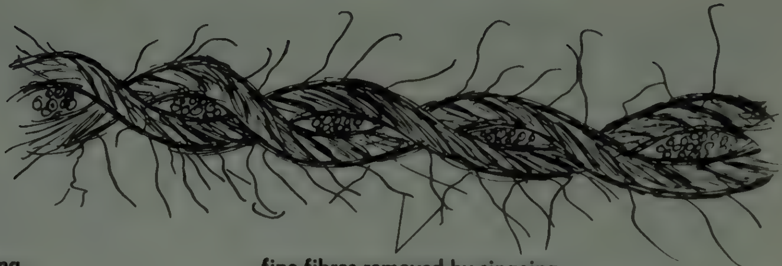


BASIC TREATMENT
chemical solvents
oil, grease,
sweat residue, etc.
drum
solid particles
(dust, etc.)
fibre
centrifuging
and
drying
fibre
solvent
dissolves
the sticky
matter
AFTER-TREATMENT
remaining adherent solid
particles are dislodged
fibre
dirty
solvent
distillation
(purification of the solvent)
AFTER-TREATMENT
water,
soap, etc.
drum
washing
drum
centrifuging
highly diluted acid
waste water
drum
drum
drum
rinsing,
centrifuging
waste water
rinsing,
centrifuging
acid treatment
(reviving the colours)
waste water
impregnating,
drying
heating
ironing
returned to customer

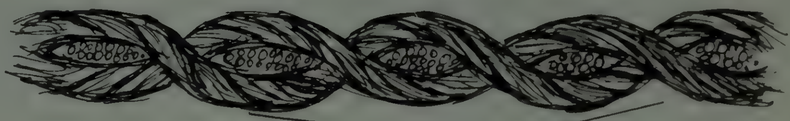
The appearance, glaze and shape-retaining properties of garments may be impaired by washing. Starching helps to restore them. This treatment consists in impregnating the garment, before ironing, with a solution of starch in water. A similar but much more comprehensive process is applied to newly manufactured fabrics. The term "finishing" or "dressing" is collectively applied to the various treatments involved. These comprise mechanical treatment and processing by chemicals to improve the glaze, shape-retaining properties, crease resistance, "feel", smoothness and drape of the material. In addition, depending on the kind of material and the purpose for which it is to be used, it can be made shrinkproof, water-repellent, supple, soft or heavy. Mechanical finishing treatments may consist in mangling, pressing, rolling, milling, shearing, calendering, raising and/or singeing (cf. page 98). Before undergoing these treatments, the material is passed through liquid baths or steam baths in which various substances (textile auxiliaries) are applied to the fibres. Solutions or suspensions in water of starch or starch derivatives, vegetable gums, glues, gelatins and mucilages improve the shape-retaining properties after the material has been dried and smoothed. In recent years, these vegetable substances have in part been superseded by more water-resistant synthetic substances, such as suitable solutions or emulsions of various synthetic resins. Oils, alcohols (e.g., glycerine), fats or tallow improve the suppleness; gypsum, kaolin, wax and binders such as albumen, glues and suitable synthetic products enhance the glaze and weight of fabrics.

Various designations, such as "non iron" or "rapid iron", are applied to textile materials which have been treated by chemical impregnation with synthetic resins and/or cellulose derivatives, usually followed by mechanical treatment applied in conjunction with heat. In particular, the tendency to crease and the swelling capacity of the fabric are reduced. White fabrics may be treated with special whiteners which enhance the impression of whiteness. Many of the organic or vegetable textile auxiliaries are liable to be affected by bacterial or fungoid decay in warm and damp surroundings; this is prevented by the addition of antiseptic substances (salicyclic acid, boric acid, zinc salts, formaldehyde compounds, etc.). Such substances may additionally perform a dirt-repellent or deodorising function. The various substances incorporated into the fabric may constitute as much as 10–20 per cent of the weight of the fabric.

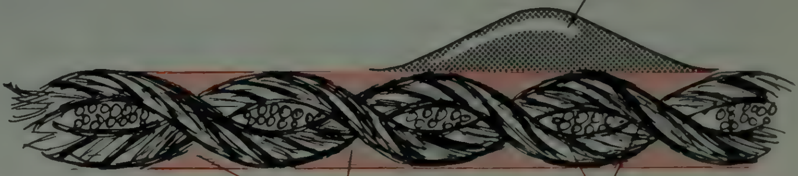
rough fabric



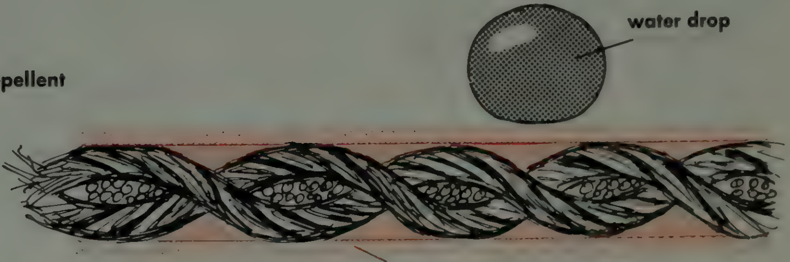
singeing, smoothing



improving the glaze



making water-repellent



silicone molecule

The letters (types) used in modern printing are made of metal consisting of approximately 70% lead, 25% antimony and 5% tin. On the front of each letter is a groove, called the "nick" (Fig. 1b), which provides a means of distinguishing the various kinds of type from one another and also enables the compositor to check the type he has set.

Printing type is available in a wide variety of designs and styles (often named after the designer who first conceived them) and in a number of different sizes (e.g., nonpareil, cicero, etc.; in English terminology, however, the type size is usually indicated as a certain number of points, ranging from 6-point, which is the smallest size used in commercial printing, to 48-point) and thicknesses (light, medium, bold, extra bold, etc.). Those parts of type matter which remain blank are filled up with what are called furniture and spacing materials, e.g., spaces between words, "reglets" (for increasing the spacing between the lines of type), and "pieces of furniture" for forming relatively large blank spaces.

The compositor works at a composing frame (Fig. 2) which has a sloping top carrying the case. The case comprises about 130 compartments of various sizes in which the letters (types) are accommodated. In the upper part of the case are the capitals, while the small letters are disposed in the lower part. For this reason capitals and small letters are referred to as "upper case" and "lower case" respectively. The latter are so arranged that the most frequently used letters come most conveniently to hand. In addition, the case contains a range of other types, such as figures, punctuation marks, and spaces for insertion between words.

In his left hand the compositor holds his composing stick (or setting stick), adjusted to the required length of the line of type. With his right hand he builds up the line, letter by letter, the letters being held in position by the thumb of the left hand. By feeling the nicks the compositor can tell that the letters are set the right way up. Spaces are inserted between the words. When the complete line has been set, these spaces are somewhat increased or reduced so as to make all the lines equal in length. When the composing stick is full, the lines of type are transferred to a so-called galley. In this way a whole page (or column) of type is assembled and is then tied up with a strong cord (Fig. 4). A proof, called a "galley proof", is then printed on a hand press and is corrected to remove any misprints. For the actual printing operation the type matter is gripped in a rectangular frame, called a chase, with the aid of metal quoins and furniture (Fig. 5). After printing has been completed, the forme (the page of type matter) is broken up and the letters are returned to their compartments in the compositor's case (Fig. 6).

In modern printing practice, hand composition is little used except for so-called jobbing work (letterheads, leaflets, etc.) and for difficult scientific typesetting (formulae, tables, etc.). For other kinds of work (books, newspapers, magazines) mechanical composition is almost exclusively employed (see page 122).

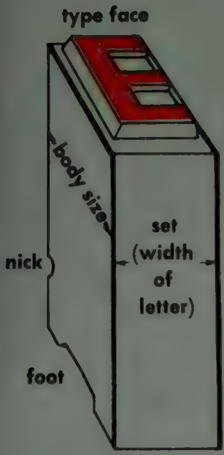


Fig. 1a

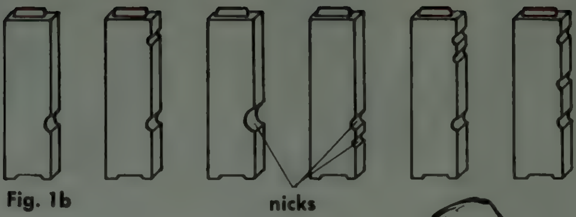


Fig. 1b

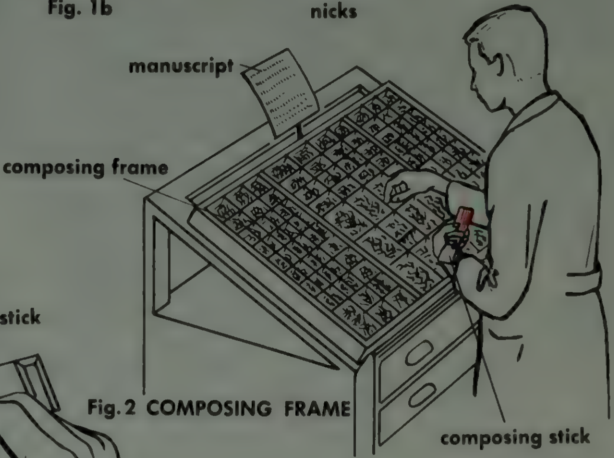


Fig. 2 COMPOSING FRAME

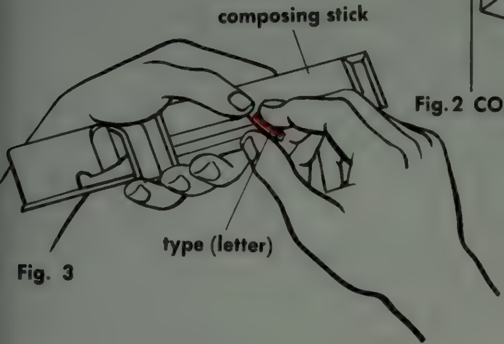


Fig. 3

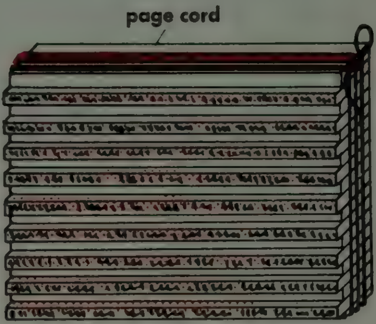


Fig. 4 TIED-UP TYPE MATTER

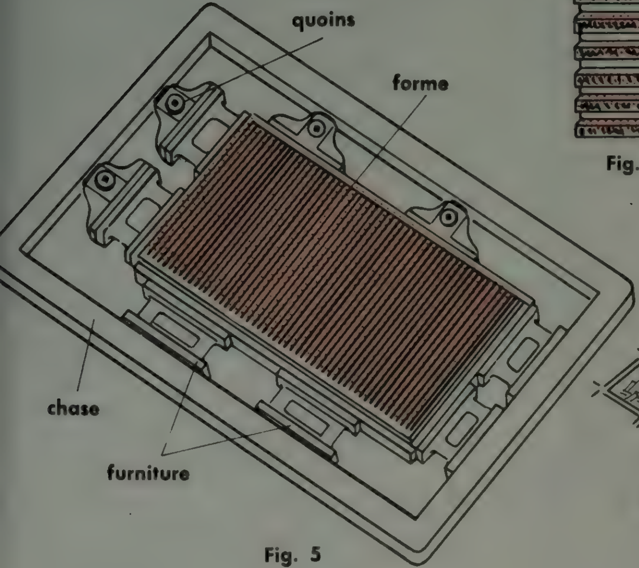


Fig. 5

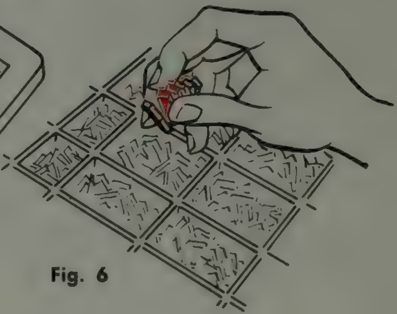


Fig. 6

Type-setting by hand. Imprimeries Réunies, Lausanne, Switzerland
Photo CIAG

TYPESETTING:
MECHANICAL COMPOSITION, COMPOSING MACHINES

The advantages of mechanical composition over hand composition are the much greater speed of setting and the use of newly cast clear type which does not have to be broken up after printing but is simply melted for recasting.

In *line casting machines* (e.g., Linotype) the typesetting—i.e., the setting up of the matrices side by side—and the casting of the lines of type are performed in the same machine. The operator sits at the machine and works a keyboard resembling that of a typewriter. Each time a key is pressed, a matrix is released from the magazine (Fig. 3) and slides on to a moving belt which conveys it to the assembler (Fig. 1). Special matrices, which are kept available in boxes on the right of the keyboard, can be inserted by hand, as required. When a complete line of matrices has been assembled, the operator presses a lever and the entire line is raised and transferred to the casting mechanism. Here the spacebands (sliding wedges) between the words are pushed home so that the line fills out to its full length. Then molten metal is pumped from the melting pot (Fig. 2) into the matrices, so that the line of type is cast (Fig. 4). The "slug" (line of cast type) is ejected and trimmed by knives. The line of matrices is transferred to grooves and they are automatically returned to their appropriate channels in the magazine. The matrices are provided with teeth cut in particular combinations; in the distributing mechanism these teeth engage with corresponding grooves, so that each matrix drops into its proper channel. The matrices are thus in constant circulation while the machine is in operation.

The Lino–Quick system operates on the same principle as the Linotype, but the operator's keyboard punches holes in a paper tape which is fed into a casting machine and controls the type-casting. Another device of this kind is the Teletypesetter. The holes are punched in the tape by a perforating machine by electrical impulses received over telephone wires, the operator being many miles away.

In the *Monotype* system the typesetting and the type-casting operations are performed by separate machines. A keyboard machine (Fig. 5) punches holes (according to a certain code) in a paper tape (Fig. 6). The tape is then fed into an automatic composition caster (Fig. 7). The letters are cast as single types (Fig. 8): this makes for greater convenience of correction, as the letters can be replaced individually, whereas with Linotype a whole line has to be recast if there is an error in it.

There are various kinds of *filmsetting machine*. One of these is the Monophoto (Fig. 9). This machine is controlled by a paper tape perforated on a Monotype keyboard. The tape controls the movements of a matrix case comprising a number of transparent plastic matrices through which a beam of light photographs each required letter in turn. When the line is completed, the strip of film is advanced on a revolving drum.

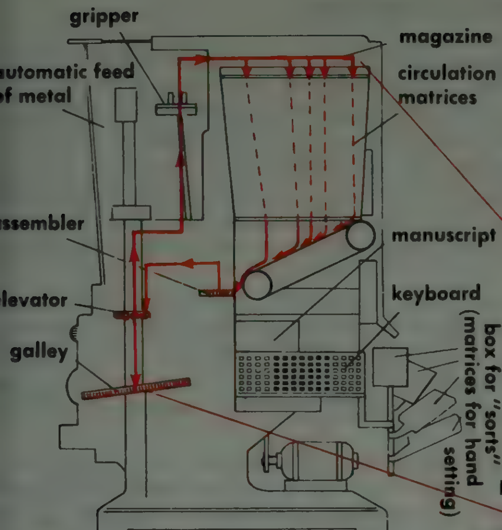


Fig. 1 LINE CASTING MACHINE

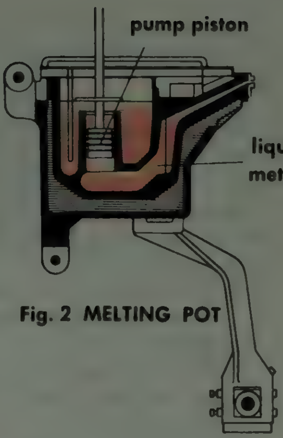


Fig. 2 MELTING POT

Fig. 3 MATRIX

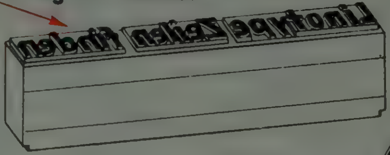


Fig. 4

Fig. 8

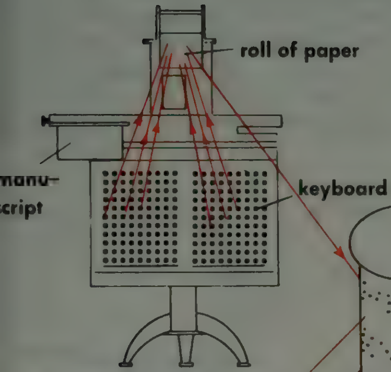


Fig. 5 MONOTYPE SETTING MACHINE

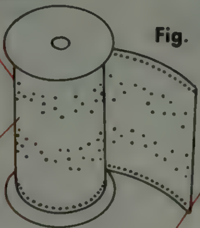


Fig. 6

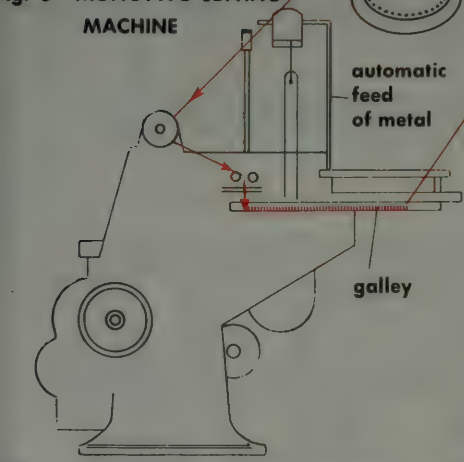


Fig. 7 MONOTYPE CASTING MACHINE

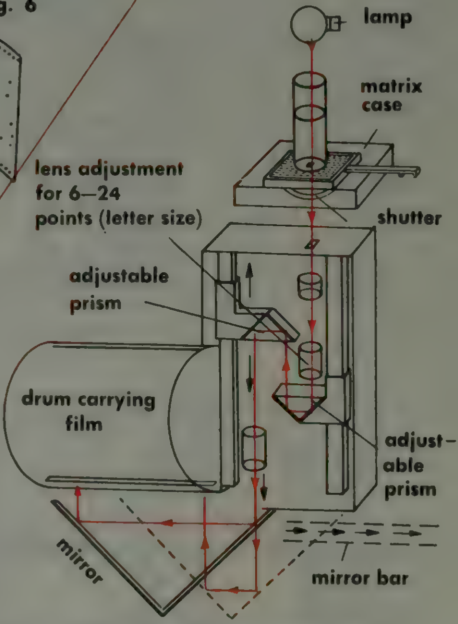


Fig. 9 FILMSETTING MACHINE (*Monophoto*)

In letterpress printing the image is printed direct from type or other relief surfaces (the term "relief printing" is sometimes used to denote this more particularly). Early printing was done on hand-operated presses, as illustrated in Fig. 1. A press of this kind comprised a screw spindle operated by a long bar, causing the platen to be pressed down on to the paper. This simple method is still used by artists for obtaining prints from woodcuts, etc. The same principle is applied in the modern sheet-fed automatic platen press (Fig. 2). Inking the printing plate in a modern machine calls for a complex inking mechanism comprising a number of rollers for evenly distributing the ink.

In the cylinder press an impression cylinder takes the place of the flat platen. The printing plate is mounted on a flat bed which is thrust to and fro under the cylinder. A letterpress machine of this kind is illustrated in Fig. 3. The sheets enwrap the impression cylinder, are held in position by grippers, and are pulled between the cylinder and the plate. With this printing system it is possible to operate at high speeds and use larger plates.

For still higher speeds, more particularly for newspaper and magazine production, web-fed rotary presses are used (Fig. 5). The curved printing plates for these machines are cast in metal by a process called stereotyping. A matrix (Fig. 4) is a mould which is formed by making an imprint of the printing forme (type and engravings) in a material such as papier-mache. A mixture consisting of 75–82 parts of lead, 15–20 parts of antimony, and 3–5 parts of tin is poured into the matrix. Up to about twenty castings (stereotypes) can be obtained from one and the same matrix. The metal printing plate produced in this way is placed round a cylinder (instead of being placed on a flat-bed machine as in Fig. 2). The continuous web of paper coming from large rolls passes between the impression cylinder and the plate cylinder, both of which rotate. Web-fed rotary letterpress newspaper machines can attain very high production rates. Sheet-fed rotary letterpress machines are used mainly for producing catalogues, magazines, etc.

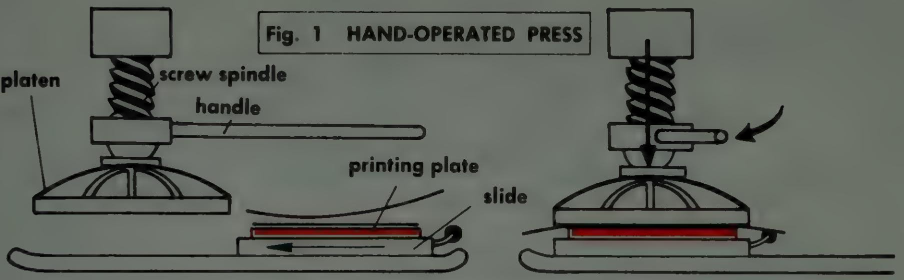


Fig. 1 HAND-OPERATED PRESS

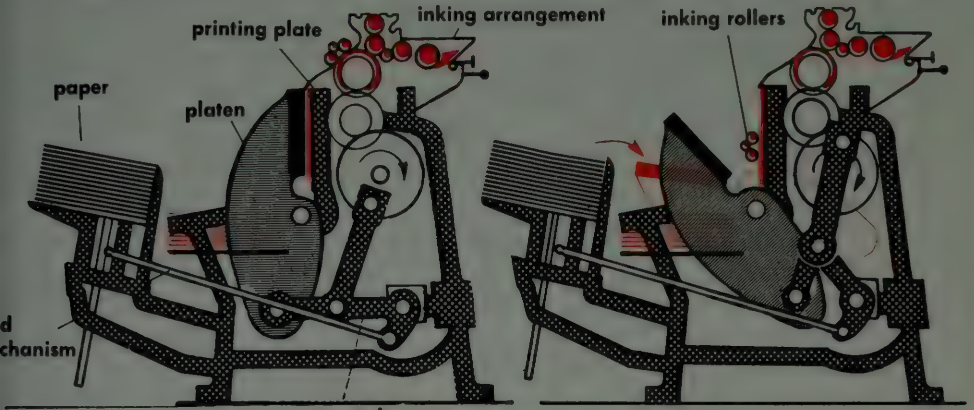


Fig. 2 AUTOMATIC PLATEN PRESS

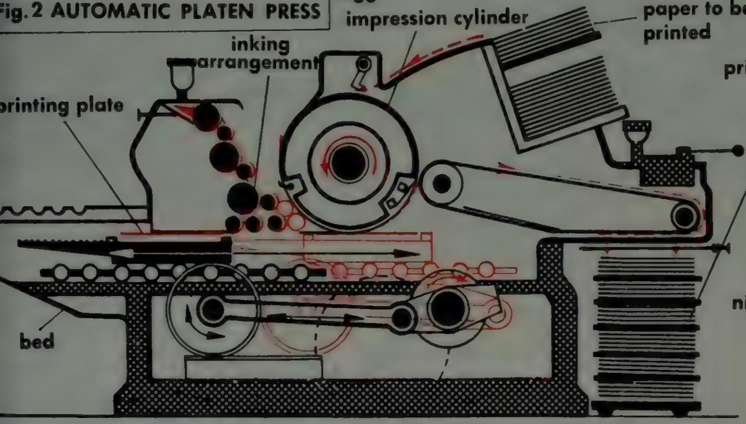


Fig. 3 CYLINDER PRESS

Fig. 4

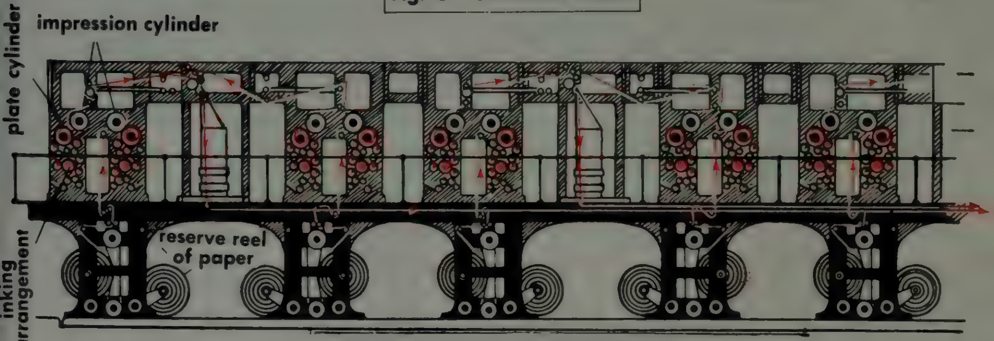
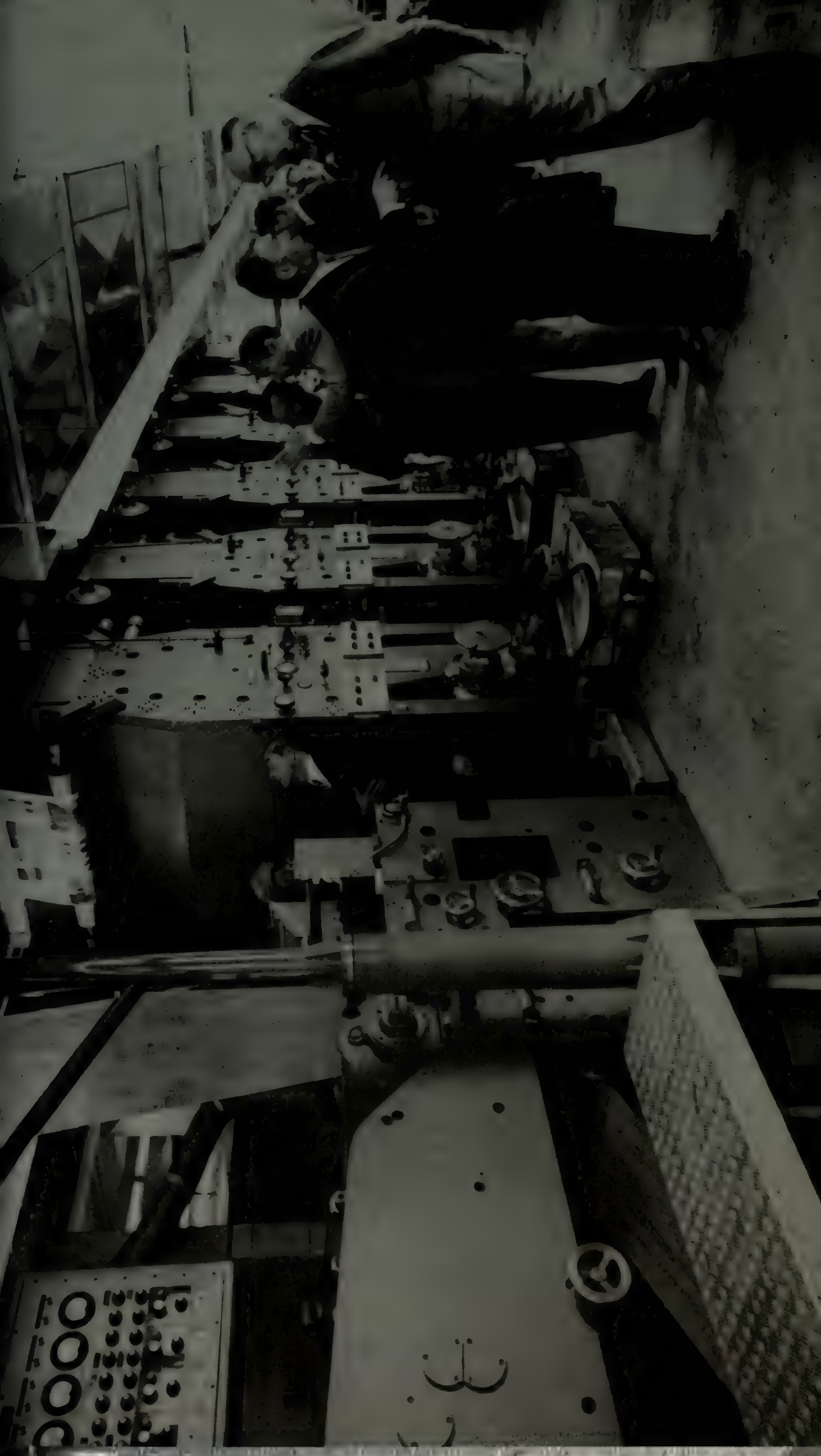


Fig. 5 WEB-FED ROTARY PRESS

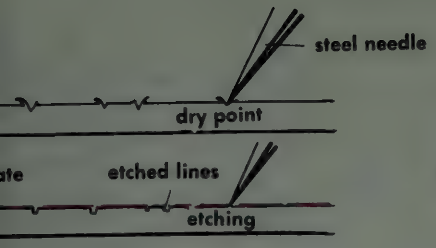
High-definition rotary printing-presses. Imprimeries Reunies, Lausanne, Switzerland
Photo CIAG



In relief printing (letterpress printing: see page 124) the lines or points to be printed are raised above the general level of the printing plate. On the other hand, in intaglio printing they are below the surface of the plate. The oldest intaglio printing process is line engraving, in which the design is produced by lines cut in the surface of a copper plate. To obtain prints, ink is applied to the face of the plate and fills the lines, any excess ink being wiped off; prints are made by laying sheets of paper on the inked plate and applying pressure by means of a roller (Fig. 2). Etching is a development of this technique; a polished copper plate is covered with so-called etching ground (a thin coating of a varnish-like substance) in which the design is scratched with a sharp needle, so that the coating is removed and the bare copper thus exposed along these lines. The plate is then treated with sulphuric acid or iron chloride, which eats away some of the exposed copper by chemical action and thus etches the design into the plate (Fig. 1b). In the allied techniques called dry point no etching chemicals are used, the design being scratched direct in the plate with a steel needle (Fig. 1a). It is therefore really a variety of engraving, but differs from it in technical details. In all these processes, therefore, the lines forming the design are below the general surface of the plate. A thick and absorbent paper is used for printing from engraved or etched plates, so that the paper will to some extent allow itself to be pressed into the lines in the surface of the plate. These centuries-old techniques are still used by artists. Commercial printing makes use of cylinder presses and rotary presses, but the actual printing operation is based on the same principle as that described above. The printing forme is almost invariably produced by photographic processes. Intaglio printing can, like relief printing, be used for a wide variety of purposes ranging from fine line drawings to colour photographs. Half-tone illustrations, such as photographs, have to be converted into a printable form by breaking up the image into small dots which form the printing surface after etching. The dots are produced by making a photographic exposure through a glass screen divided into small squares by intersecting dark lines. In this way a dot image on the photographic plate is obtained. For printing on cheap coarse paper (such as newsprint) a screen with 60 lines to the inch is used; plates made with finer screens must be printed on a better-quality coated paper. From the negative thus obtained a print is made on copper plate coated with a photo-sensitive solution. The plate is then etched, whereby the non-printing areas are removed, leaving the dots standing in relief. The technique of producing pictures by this method is known as the half-tone process. It comes under the more general heading of photo-engraving. For intaglio printing the dots are not raised above, but are "wells" depressed below the non-printing areas; the latter have no ink on them, the ink being concentrated in the wells, which are deeper or shallower, depending on the tonal value of the original. These plates are produced by a method similar in principle to that for making relief printing plates, but using a different type of screen and different technical processes.

Intaglio printing is done on presses generally similar to those used for letterpress printing, such as the sheet-fed machine (used more particularly for high-quality work) illustrated in Fig. 4, except that in intaglio printing the printing plate is always wrapped round a cylinder (instead of a flat plate, as in the cylinder press described on page 124). The plate is inked with a thick soft ink which fills the "intaglio" depressions, and any excess ink is scraped off by a steel blade (called a "doctor"). There are many different inking systems; Fig. 5 shows one of these. A web-fed rotary press is illustrated in Fig. 6.

The intaglio printing process described here is known as "gravure printing" or "photogravure". The term "rotogravure" is used more particularly for high-speed printing on web-fed machines, such as that in Fig. 6, which produces catalogues, newsprint supplements, packaging materials, etc.



. 1 ETCHING

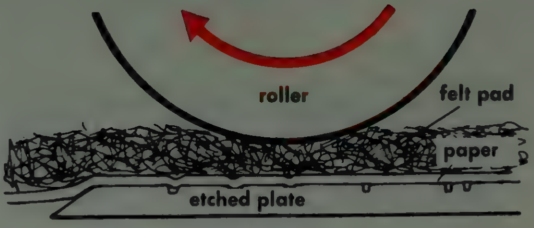


Fig. 2 PRINTING

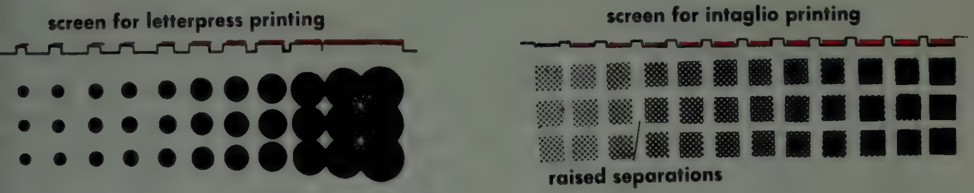


Fig. 3 SCREEN

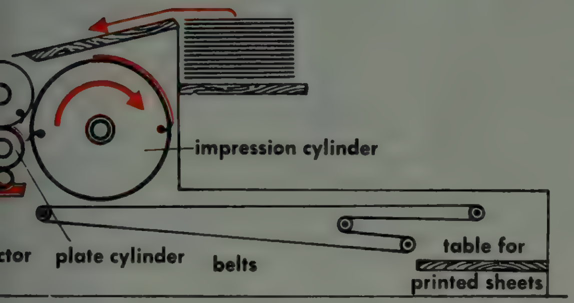


Fig. 4 SHEET-FED PRINTING PRESS

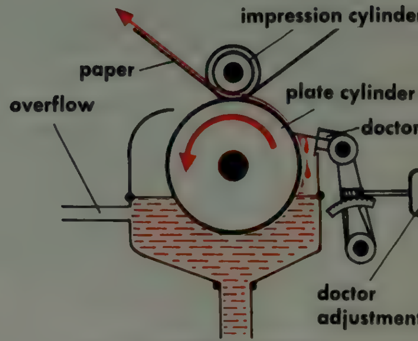


Fig. 5 INKING SYSTEM

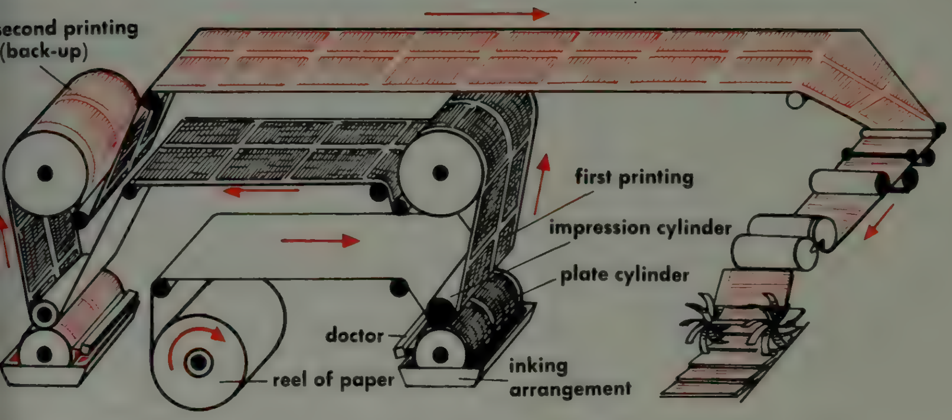
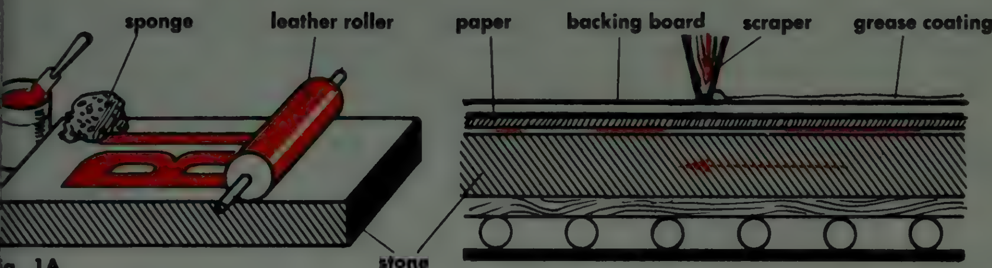


Fig. 6 WEB-FED ROTARY PRESS

PLANOGRAPHIC PRINTING PROCESSES: LITHOGRAPHY AND OFFSET PRINTING

In planographic processes the matter to be printed is not raised above the surface of the plate (as in letterpress or "relief" printing) nor sunk below the surface (as in intaglio printing), but remains on the surface. Lithography is based on the principle that grease and water are mutually repellent. Ink is applied to grease-treated printing areas, while the non-printing parts, which absorb and hold water, reject the greasy ink. The process was invented by Senefelder, a German, in 1798. The drawing is made with a greasy medium on a special kind of stone, which is then lightly etched. Next, the stone is treated with gum arabic and the drawing is washed off with turpentine. Finally, the stone is coated with water and inked (Fig. 1a). Printing is done by applying sliding pressure exerted by a wooden scraper. The stone is covered with dampened paper and a backing board. Then the stone (on the bed of the press) is moved horizontally (Fig. 1b).

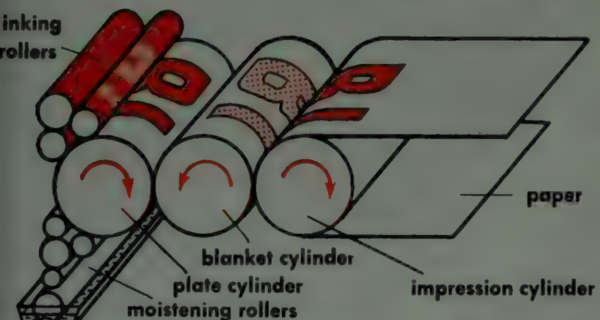
A similar technique can be applied to metal plates instead of stone. These have the advantage that they can be wrapped around the cylinder of a printing press and can thus be used for printing by rotary methods. As a rule, the image is not printed direct from the plate on to the paper, but is first transferred to an intermediate rubber cylinder (blanket cylinder) which then transfers (offsets) the image to the paper. This procedure has the great advantage that, because of the flexibility of the rubber cylinder, the image can be transferred not only to paper, but also to various other materials, rough or smooth. The offset printing press thus comprises three rotating cylinders (Fig. 2): the plate cylinder (carrying the printing plate), the blanket cylinder (covered by a sheet of rubber), and the impression cylinder (which presses the paper against the blanket cylinder). Moistening rollers apply a film of water to the plate, which then comes into contact with inking rollers (Fig. 3). The ink is rejected by the water-holding areas but is accepted by the greasy image. The process is used for the printing of copy (text) as well as pictures and is widely used for the production of books and magazines. Fig. 5 is an illustration of a four-colour offset printing press comprising two plate cylinders, each with four sets of inking rollers (one set for each colour) and each printing on one side of the web of paper. The printing plates for offset lithography are produced by photographic processes, the metal plate (zinc or aluminium) being given a sensitised coating. The exposed plate then undergoes various treatments (developing, washing, etching, etc.) before it is ready for use as a printing plate. The offset lithography process is also known as photo-lithography.



ig. 1A

Fig. 1B

Fig. 1 LITHOGRAPHY



. 2 PRINCIPLE OF OFFSET PRINTING

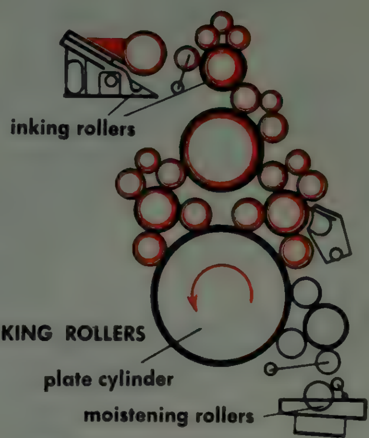
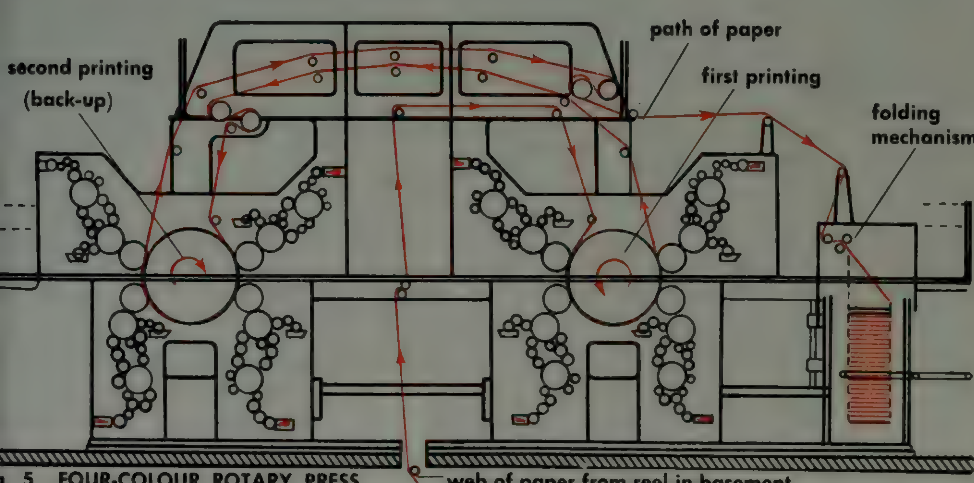


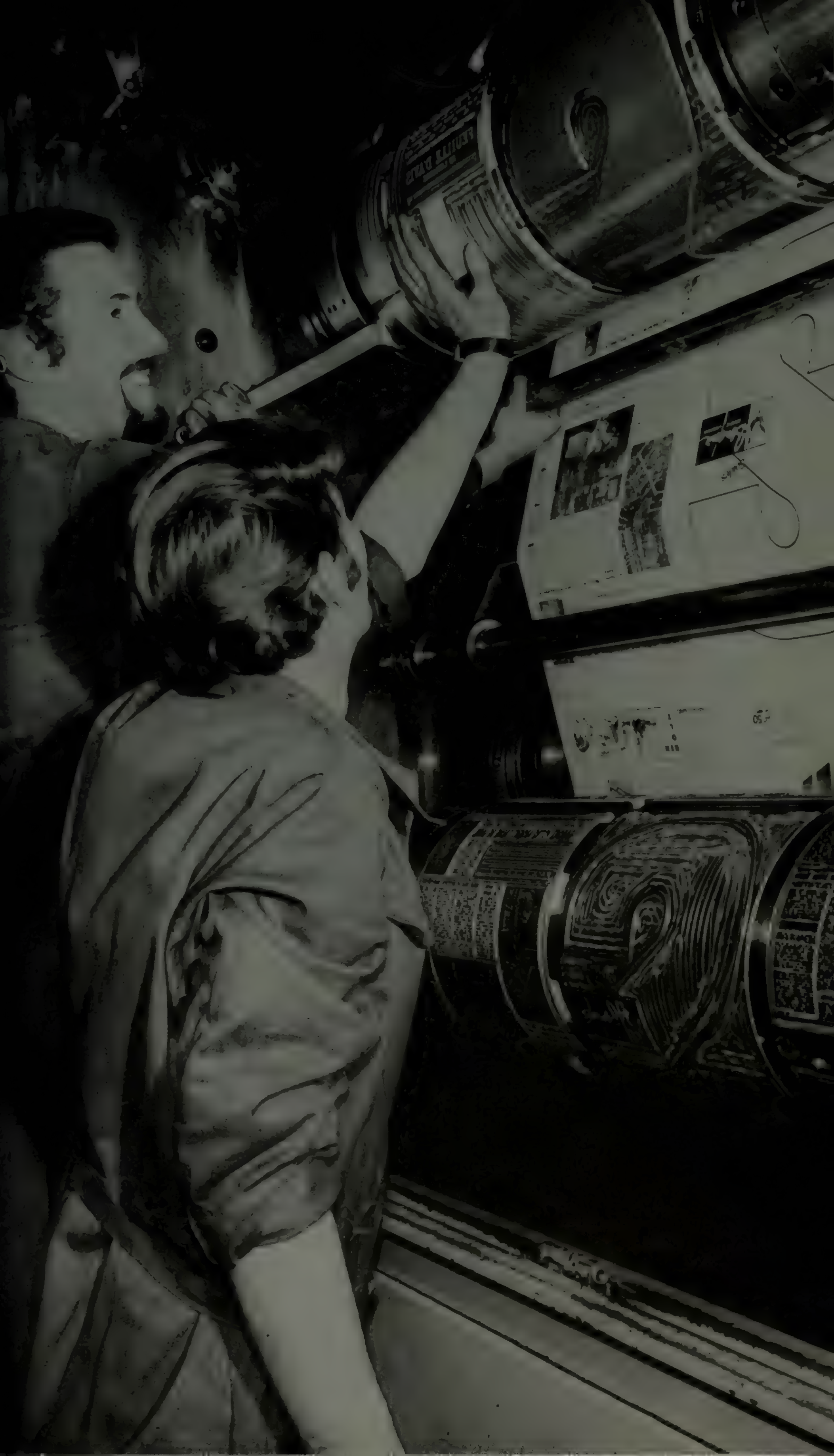
Fig. 3 INKING ROLLERS

ig. 4 SHEET-FED TWO-COLOUR PRINTING PRESS

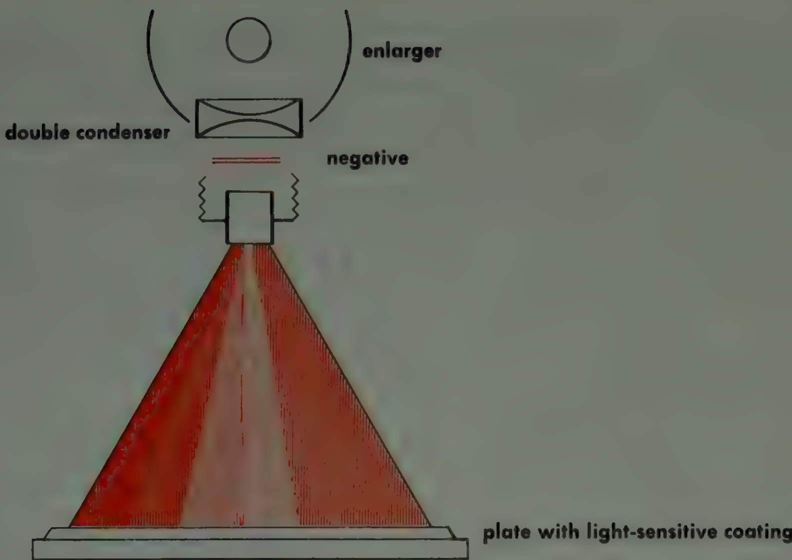


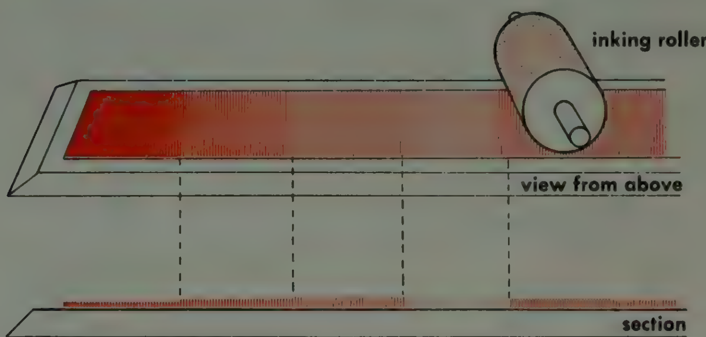
. 5 FOUR-COLOUR ROTARY PRESS

Attaching the plate to the cylinder.
Imprimeries Réunies, Lausanne, Switzerland
Photo CIAG

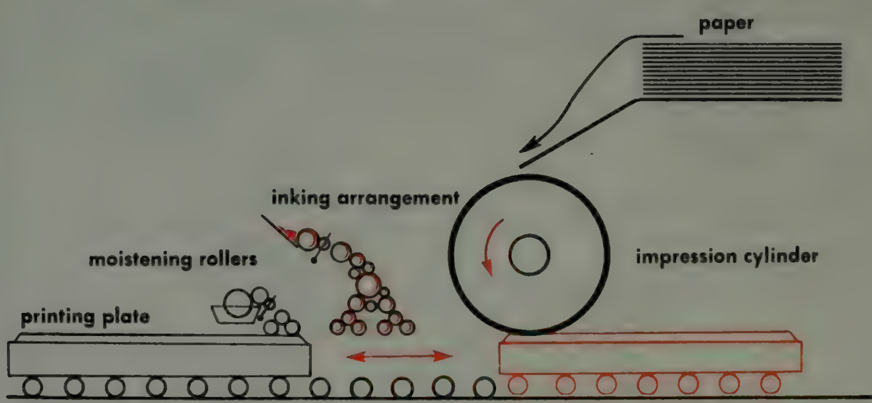


The collotype, or photogelatin, printing process differs from letterpress, offset and photogravure in that the image is not broken up into dots by means of a half-tone screen. This enables it to make more closely accurate reproductions than any of the other processes. It resembles photolithography (see page 130) in that it is based on the fact that water and grease repel each other. The non-printing areas of the plate hold the moisture and reject the greasy printing ink, while the relatively dry printing areas accept the ink. A glass or aluminium plate is coated with a light-sensitive solution containing gelatin and potassium bichromate. The image of a photographic negative is then projected on to this plate. Where the light passes through the transparent areas of the negative, the gelatin is hardened; darker areas prevent hardening to a greater or less degree. Next, the plate is soaked in glycerin, which is absorbed chiefly by the soft (non-printing) gelatin areas. The plate is exposed to a moist atmosphere, and the soft areas absorb moisture. When the plate is inked, these areas repel the greasy ink, whereas the hardened areas, which do not absorb moisture, do accept the ink. Printing is done on flat-bed presses or on cylinder presses.





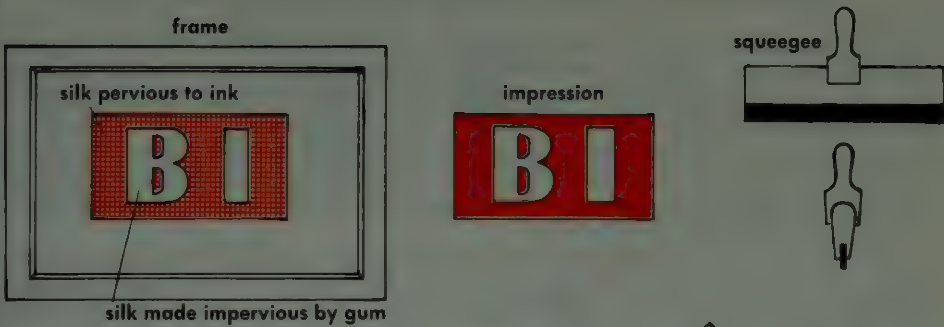
light-sensitive coating
developed and moistened



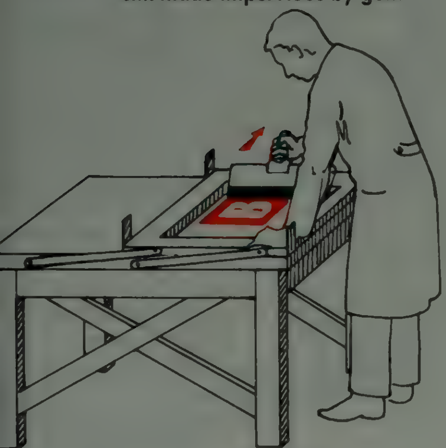
principle of planographic printing

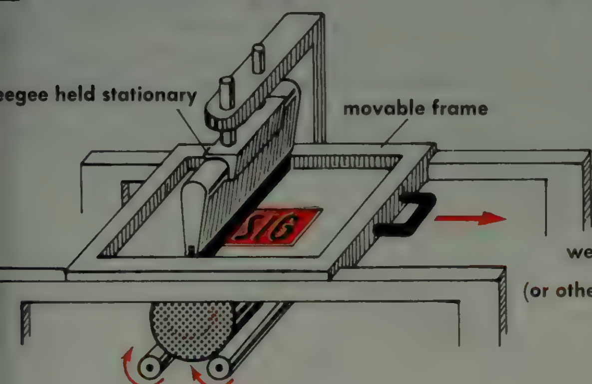
SILK SCREEN PRINTING

This is essentially a stencil printing method, the stencil being formed by bolting silk stretched over a wooden or metal frame. For fine detail a fine-meshed silk will be used. Steel wire gauze or plastic screens are sometimes used for special purposes. The design can be applied directly to the screen by painting it with a greasy medium (paint, ink, etc.). A water-soluble gum is then applied to the screen and closes the pores of the silk. However, the greasy areas reject the gum. The greasy paint is then washed away with a solvent (e.g., turpentine), so that the corresponding areas of the silk become pervious to ink. The screen is then placed on the surface to be decorated and ink is applied to the screen by means of a rubber squeegee. Some of the ink soaks through the pervious areas of the silk and is printed on to the surface (paper, metal, wood, glass, rubber, textile fabric, etc.). The screen-printing of textiles is carried out on long tables, the screen being moved along step by step to successive positions (as shown in the bottom illustration). To obtain fine detail in the pattern, it is necessary to produce the stencil by a photographic process. This is done by applying a light-sensitive coating to the screen. The image is developed by washing: this removes the coating from those areas where it has not been hardened by the action of light. The silk screen process is very versatile with regard to the wide range of materials to which it can be applied, but it is not suitable for reproducing fine detail and gradations of tone. Simple designs can be made by sticking paper, celluloid or metal patterns to the screen. The screen can be wrapped round a cylinder for rotary printing (see illustration).



frame
silk pervious to ink
silk made impervious by gum
impression
squeegee





eegee held stationary
movable frame

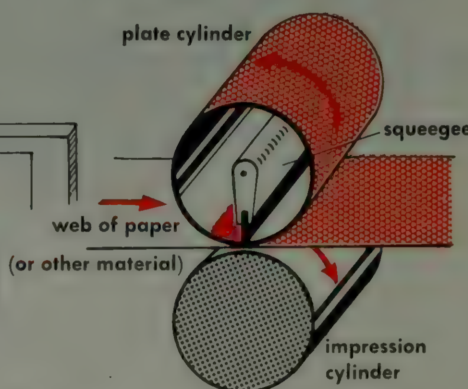
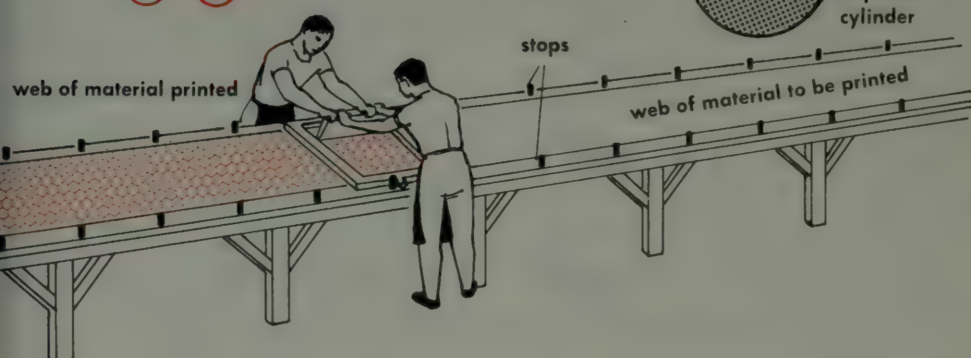


plate cylinder
squeegee
web of paper
(or other material)
impression
cylinder



web of material printed
stops
web of material to be printed

COPYING AND DUPLICATING (XEROGRAPHIC METHOD)

The xerographic process was evolved by Carlson in America in the nineteen-thirties. He used a photoconductive semi-conductor, a material which conducts electricity on exposure to light but behaves as an insulator in the dark. This material was applied in layers to a conducting plate. The semi-conductor was electrostatically charged in the dark, and a pattern was then projected on to it. In the illuminated areas of the design the electric charge was dissipated, but the unilluminated areas retained as residual charge which was made "visible" by dusting the plate with a suitable powder (e.g., rosin powder).

In the present-day application of this principle a colouring substance is applied to an electric charge pattern formed in a semi-conductor, which nowadays consists of coatings of selenium, selenium/arsenic or selenium/tellurium. What is known as the cascade method of development has come into widespread use in modern xerographic reproduction technique: A very fine synthetic powder (called the toner) with special frictional electric properties are mixed with small steel and quartz balls provided with a coating of special plastic. As soon as this mixture is set in motion, the balls and the synthetic powder induce electric charges in each other, the charge acquired by the powder being of opposite sign to that of the semi-conductor. If a mixture of this kind is moved about on the coated plate, particles at first adhering to the carrier balls are detached from these and remain adhering to the more highly charged areas of the plate. The powder thus makes the pattern of electric charges in the plate "visible". As a result of the development process, which takes only a few seconds, a complete powder image emerges on the plate. This image is transferred to paper likewise by means of electrostatic forces. The paper is simply laid upon the powder image and is electrically charged from the back. This causes the powder to adhere to the paper. To fix the powder, heat is applied, which causes the thermoplastic powder particles to melt and thus remain permanently adhering to the paper. In other methods the powder is sprayed with a solvent or is fixed by a varnishing treatment. Alternatively, some methods print the powder image on to paper coated with plastic and zinc oxide, the image being directly fixed by the semi-conductive zinc oxide coating. Xerographic processes are used not only for copying and multiplying but also, for example, for obtaining rapid printing-out of results from data-processing equipment (cf. vol. I, p. 334 *et seq.*). The electric image is produced by electrodes in printing heads controlled by impulses emitted by the equipment, the image then being made visible and fixed in the manner described.

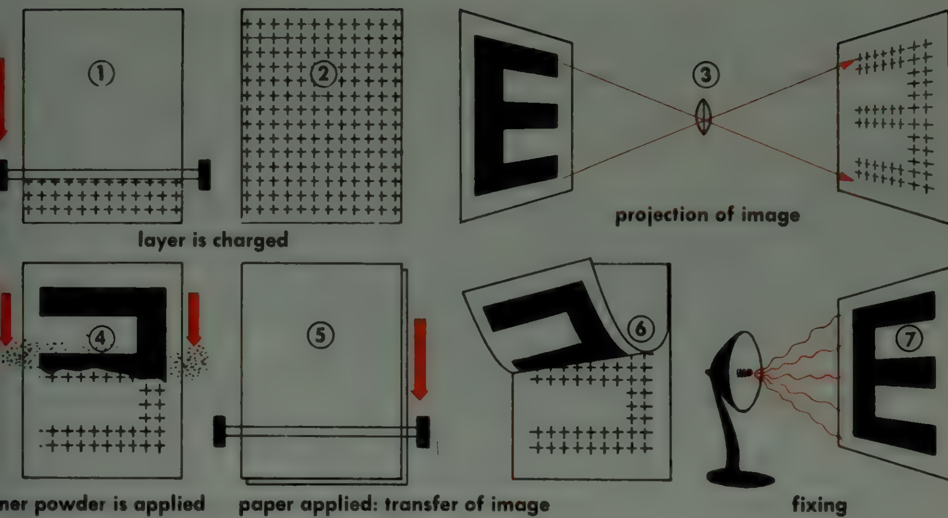
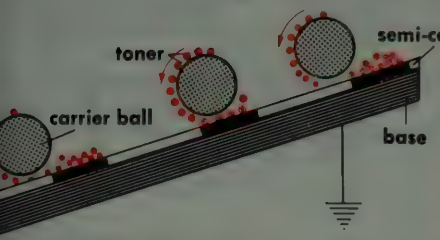


Fig. 1 PRINCIPLE OF XEROGRAPHIC PROCESS



2 PRINCIPLE OF CASCADE DEVELOPMENT

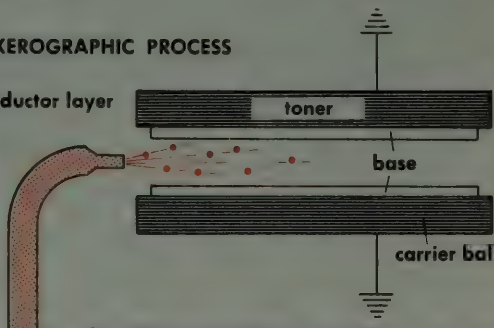


Fig. 3 DEVELOPMENT BY DUSTING WITH POWDER

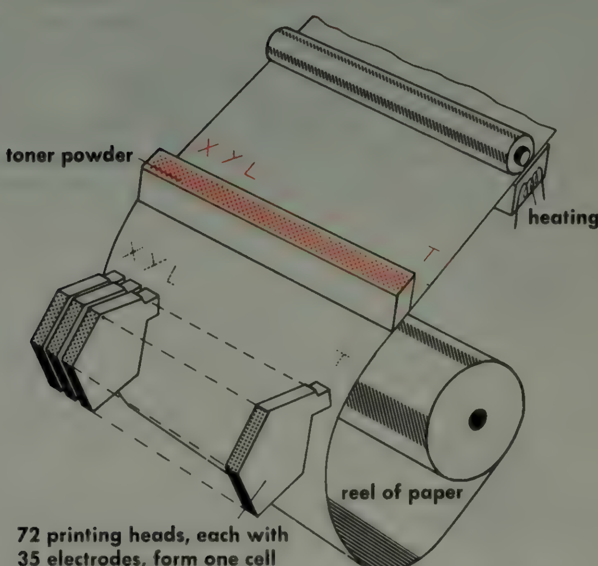


Fig. 4 HIGH-SPEED PRINTING EQUIPMENT (BORROUGHS)

The sheets arriving from the press are folded into 8-, 16-, 32- or 64-page sections (known as signatures). First, however, the sheets arriving in the bookbinding works are accurately placed in stacks and are then cut to the correct size by means of a *cutting machine* (or guillotine). The sheets are then passed to the *folding machine* in which the sheets are folded halfway down their length a number of times until the correct page size for the book and the correct sequence of page numbering is obtained.

The folded signatures are placed in the receiving hoppers of the *collating machine* (or gatherer). As many hoppers are needed as there are signatures in the book and the signatures are arranged consecutively in the hoppers. From each hopper one signature at a time is deposited on a travelling belt, which then moves it along to the next hopper, where the next signature is deposited on the first, and so on. Each cycle of the machine delivers one complete gathered book.

The gathered signatures, or sections, are then ready for sewing. This operation is done on a *book sewing machine*, nowadays usually a semi-automatic or fully automatic machine. The operator places the sections over a saddle from which they are moved automatically to the sewing position or, in the fully automatic machine, merely feeds them into a hopper. For better joining the pages together, book back glueing is applied, i.e., the backs of the sewn sections are passed over rollers which put glue on them. When the glue has dried, the treatments known as "smashing" or "nipping" are applied. Nipping is pressure applied to the back only and is usually sufficient for novels, whereas larger books are subjected to smashing (all-over compression applied to the book block). Trimming the book blocks is done on a *three-knife cutting machine* that trims all three sides.

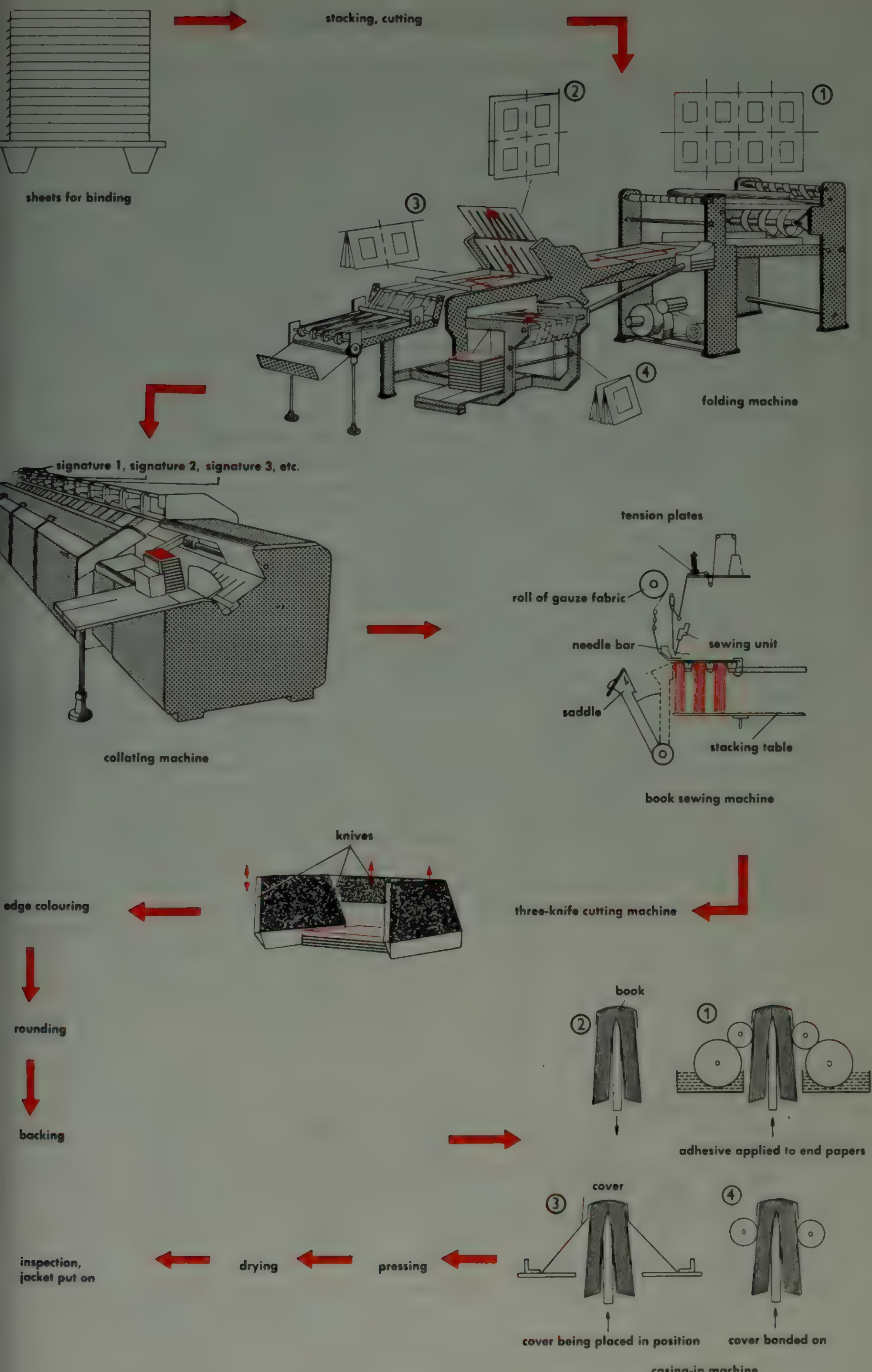
For cheaper books ("paper-backs") threadless or unsewn binding[1] is now widely used. In this system the backs of the sections are cut off, so that the book block consists of single leaves. A strip of gauze linen and glue are then applied.

After trimming, the books undergo rounding (putting the round into the back), which is done on a special machine, and backing (glueing the backbone, applying the head and tail bands[2] and the paper backbone lining).

Casing-in machines fix each book into its case (or cover) and are semi-automatic or fully automatic. The book is split by an ascending blade which moves it upwards between paste rollers. These apply adhesive to the end papers. The case, which has been treated on a forming iron to achieve correct shaping of the backbone, is then placed in position over the ascending book. From the casing-in machine the books are fed to a forming and pressing machine in which the cases are firmly bonded on by the application of pressure and heat.

The cases (covers) are made entirely by mechanised processes on casemaking machines. Some of these machines, which again are semi-automatic or fully automatic, make cases from sheets of cloth already cut to size; others have the cloth fed to them in a continuous web from a roll.

1. Called perfect binding in U.S.A.
2. Foot bands in U.S.A.


sheets for binding
stacking, cutting
①
②
③
④
folding machine
signature 1, signature 2, signature 3, etc.
collating machine
tension plates
roll of gauze fabric
needle bar
sewing unit
saddle
stacking table
book sewing machine
knives
three-knife cutting machine
edge colouring
rounding
backing
book
②
①
adhesive applied to end papers
cover
③
④
cover being placed in position
cover bonded on
casing-in machine
pressing
drying
inspection,
jacket put on

The tractor is an important aid in modern farming. In most essential respects it is very similar to an automobile. A diesel engine (see page 194) transmits its power through a propeller shaft to the eight- or ten-speed gearbox (cf. page 214) and is thence transmitted to the drive wheels through the differential gear (cf. page 224). The propulsive force and therefore the tractive force (pull) developed by the tractor depends not only on the force that the engine transmits to the big rear drive wheels but also on the load on these wheels. If the axle load is too small, the wheels will slip and fail to grip the ground. To increase the load and thus press the drive wheels more firmly against the ground, ballast weights may be provided at the rear, or sometimes the tyres are filled with water instead of air. Often the attached implements drawn by the tractor are so designed as to develop an additional downward force on the rear axle of the tractor. A modern tractor is also extensively used as a power unit for driving various kinds of farm machinery and is, for this purpose, usually provided with three power take-off shafts. Two of them are mounted under the trailer coupling and rotate at 1000 r.p.m. and 540 r.p.m. respectively, power being transmitted to the attached machinery through an articulated shaft. The third power take-off shaft protrudes forward and is intended more particularly for driving an attached grass or grain cutting machine (see page 146). The speed is 1000 r.p.m. The hydraulic equipment of the tractor is also a very useful adjunct which serves not only for adjusting the positions of the towing control rods but also for hydraulic control of attached machinery.

The hydraulic system is shown schematically in Fig. 2. The pressure is generated by a hydraulic pump (e.g., a geared pump, as shown, or alternatively a piston pump or a vane pump), which forces the oil through the delivery pipe to the control valve. In the "lifting" position the slide opens the passage to the working cylinder. The piston travels slowly out of the cylinder. Through an appropriate linkage system this piston movement raises the two bottom towing control rods. If the load to be lifted is too heavy, the pressure that builds up in the delivery pipe causes the safety valve to open and thus bypass the oil to the oil supply tank. This prevents damage to the hydraulic system. In the "neutral" position on the slide closes the passage to the working cylinder, so that the piston is locked. However, since the pump continues to deliver oil, the slide simultaneously opens a passage leading the oil, via the oil filter, back to the supply tank. The oil filter is provided with a relief valve which allows the oil to flow to the tank even if the filter is clogged with dirt. In the "lowering" position the slide again opens the passage to the working cylinder. The pressure of the attached load causes the oil to flow out of this cylinder and be discharged, along with the oil delivered by the pump, via the filter to the supply tank.

(Continued)

Fig. 1 FARM TRACTOR

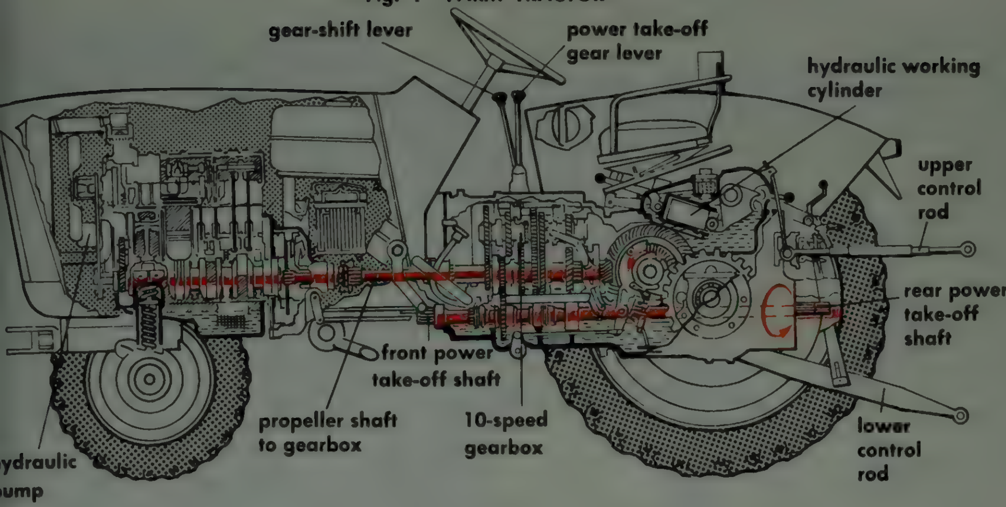
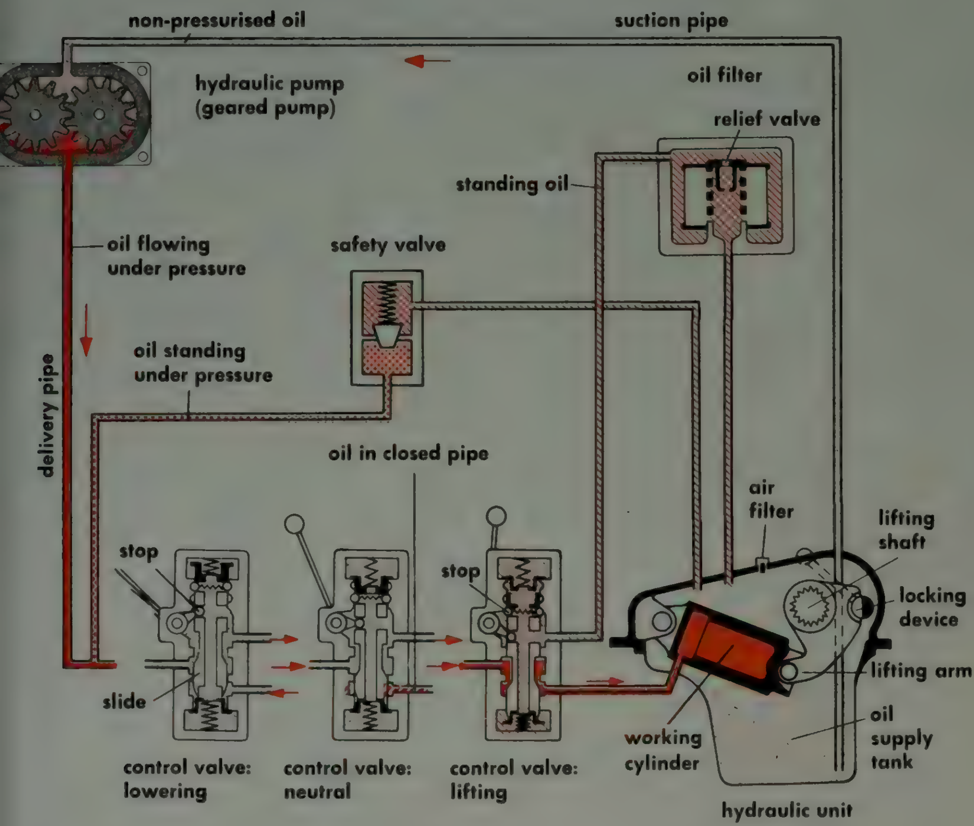


Fig. 2 HYDRAULIC SYSTEM



Farm tractor pulling a Huard plough
Photo Huard

Grass and grain cutting attachment: The third power take-off shaft (see page 142) is specially intended for driving this attachment (Fig. 1). For this purpose the shaft is fitted with a crank disc which acts through a connecting rod, toggle lever and oscillating rod and thus produces a to-and-fro motion of the cutter blades. The drive mechanism is provided with a slip clutch (built into the crank disc) as a safeguard against jamming or excessively large objects (pieces of wood, stones) getting between the blades.

Plough (Figs. 2a, b, c): Although the plough is one of the oldest agricultural implements, it has changed very little in shape and construction over the years. Only the plough tail (handles) has lost its significance in modern tractor-drawn ploughs. The coulter, attached to the plough-beam, makes a vertical cut in the ground. The ploughshare coming behind it determines the depth of the furrow. It cuts the soil horizontally and lifts it to the mould-board which turns it over. The slip heel provides bottom guidance and the landslide takes the side thrust. For completely turning the soil or for ploughing-in straw or stubble the plough is often equipped with a so-called skim coulter (Fig. 3). The ploughshare and mould-board together are called the plough bottom.

The tractor-drawn plough is attached by means of securing pins to the towing control rods, which are hydraulically controlled (see page 142).

A two-way, or reversible, plough has a right-hand and a left-hand plough bottom (usually two of each in the case of a tractor-drawn plough, as in Fig. 4a). On reaching the end of a furrow, the operator reverses the plough. He can thus plough in either direction and throw all the furrows one way. The ordinary mould-board plough (Fig. 4b) cannot do this: if all the furrows are to be thrown one way, ploughing must be done in one direction only, i.e., the return journey to the "starting" end of the field must be made with the ploughshare lifted out of the ground. Various kinds of rotary ploughs have been used in Europe for many years. A modern type, not yet widely used, is illustrated in Fig. 4c. A narrow ploughshare passes the soil to a rotor which performs the function of turning the soil, so that the tractive force needed for pulling the plough is reduced. These ploughs are claimed to be more efficient and faster than ordinary ploughs.

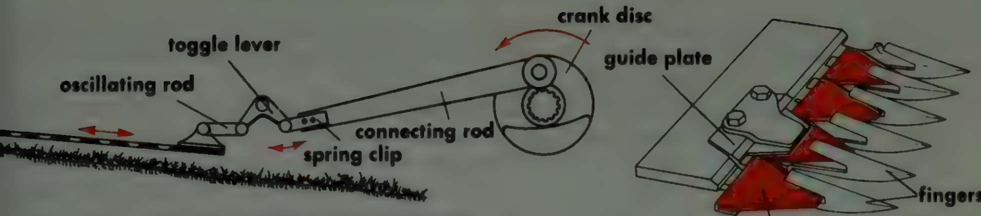


Fig. 1 GRASS CUTTING ATTACHMENT

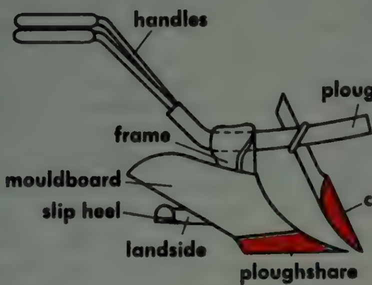


Fig. 2a PLOUGH

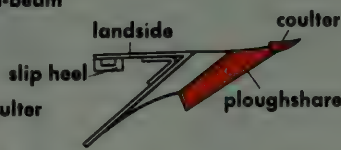


Fig. 2b PLOUGH (*top view*)

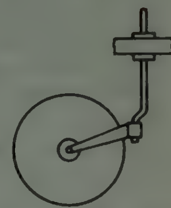


Fig. 2c DISC COULTER

Fig. 3 OPERATION OF SKIM COULTER: (*a*) *ploughing without, and* (*b*) *ploughing with skim coulter*

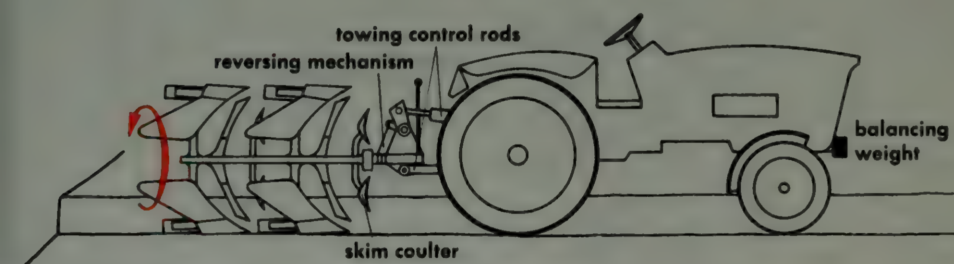


Fig. 4a TWO-WAY PLOUGH WITH SKIM COULTER

Fig. 4b TWO-SHARE PLOUGH WITH SKIM COULTER

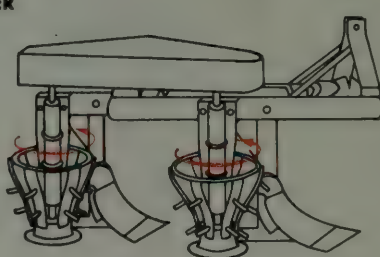
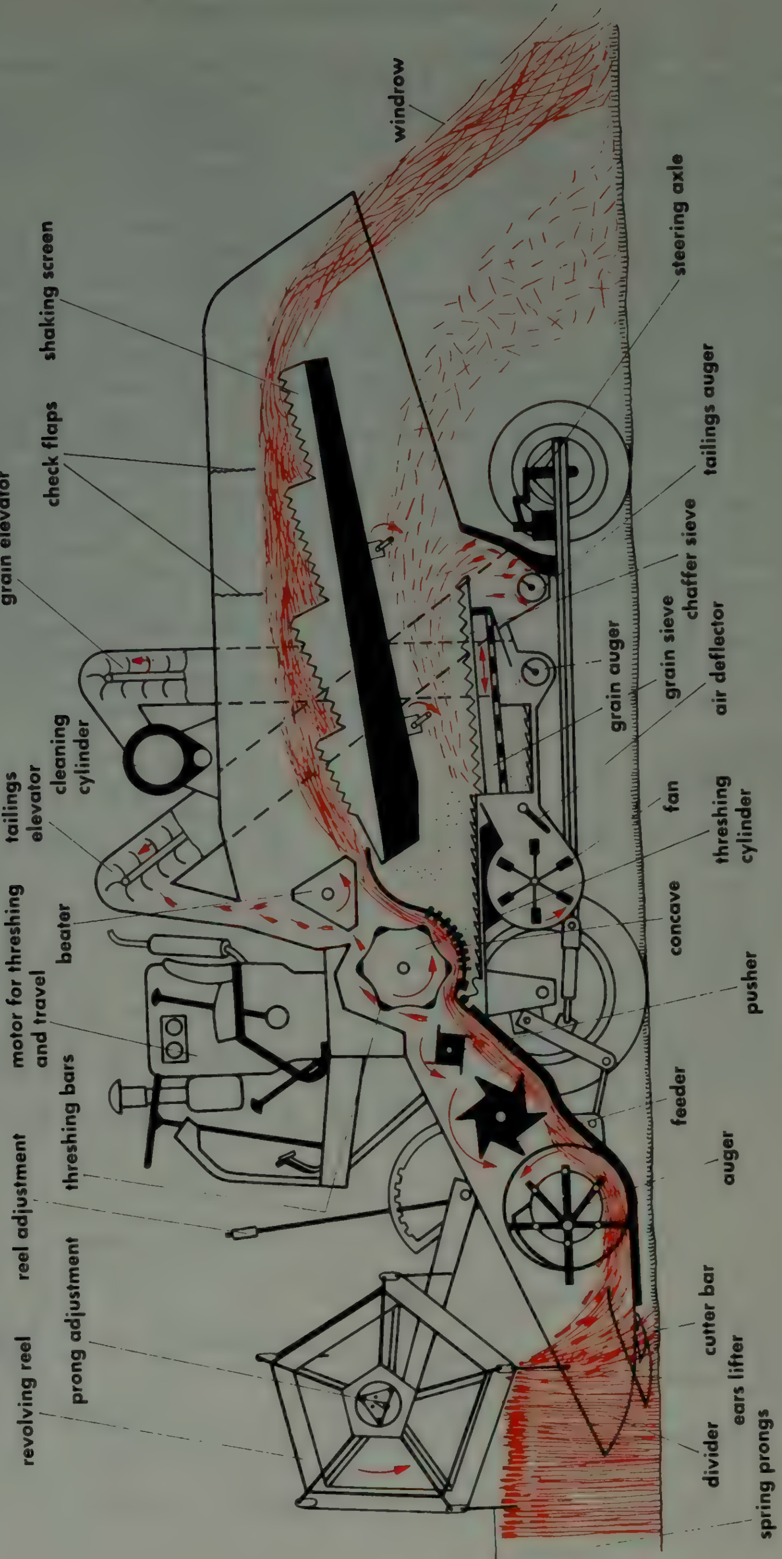


Fig. 4c ROTARY PLOUGH

COMBINE HARVESTER AND THRESHER

The combine harvester is a combination of a grain harvesting machine and a threshing machine. The grain is cut, threshed and cleaned in one operation. The machine may be self-propelled or be towed by a tractor.

The wheatstalks are cut by an oscillating knife while the revolving reel pushes them back towards the knife and auger (feed screw). Grain flattened by wind or rain is raised by the spring prongs on the reel. The cut wheat is conveyed into the machine by the auger and reaches the threshing cylinder which rubs the grain out of the heads against a "concave". The grains of wheat, together with the chaff and short fragments of straw fall through the interstices between the bars of the concave and into the cleaning "shoe". Some of the grain is carried along with the straw, is stopped by check flaps, and is shaken out of it on shaking screens on the straw rack of the machine. The straw drops out of the back of the machine and is left in a windrow for later baling, or is baled directly by a baling attachment or press, or is scattered over the ground by a fan-like straw spreader. The grain shaken out of the straw is also delivered to the cleaning shoe. In the shoe the grain is separated from the chaff and cleaned by sieves and a blast of air. The chaff and fragments of straw are thrown out from the back of the machine. The grains of wheat fall through the sieves and into the clean-grain auger (screw conveyor) which conveys them to an elevator and on into the storage tank or into bags. Any heads of wheat which fail to go through the sieves and are thrown backwards by the air blast fall short in comparison with the lighter chaff and drop into a return auger which, via an elevator, returns them to the threshing cylinder. Correct adjustment of the air blast—by throttling the intake of the fan and by altering the setting of baffles—is important in determining the degree of cleaning of the grain and the magnitude of the grain losses that occur.


revolving reel
prong adjustment
reel adjustment
threshing bars
motor for threshing and travel
beater
tailings elevator
cleaning cylinder
grain elevator
check flaps
shaking screen
windrow
steering axle
tailings auger
chaffer sieve
grain sieve
air deflector
grain auger
fan
threshing cylinder
concave
pusher
feeder
auger
cutter bar
ears lifter
divider
spring prongs

The beet harvesting machine (Fig. 1) cuts off the top of the beet (this operation is called "topping"), lifts the beet out of the ground, cleans the earth off it, and delivers it to a hopper or to a wagon travelling along with the machine. Topping is performed by a topping device which may either comprise a knife (Fig. 3) or a revolving inclined disc (Fig. 4). In the former arrangement a spiked roller moves along the ground. When it travels over a beet, the topping knife is automatically set to the correct height. The cut-off beet head with the leaves is fed by the action of the spiked roller to an elevator and is discharged into a hopper (not shown) or is deposited in a windrow beside the machine (as in Fig. 2). The lifting mechanism operates on the adjacent row of beet, already dealt with by the topping device. The guide skids keep the lifting fork at the correct pre-set working depth. The fork seizes the beet below its point of greatest girth, loosens it from the ground, and passes it to the cleaning wheel which flings the beet against a surrounding cage and there removes adhering earth. The beet is then delivered to an elevator which conveys it to the beet hopper. From time to time the contents of this hopper are emptied into a wagon. This is done by means of the discharging conveyor which forms one side of the beet hopper and which can be swung down to deliver the beet into the wagon. In addition, the hopper can (by hydraulic power) be tilted about the axis *A–A* in Fig. 1. The lifting mechanism can be swivelled by the operator, if necessary, in order to seize any beet that may be somewhat out of the row. The power for driving the beet harvesting machine is supplied through a drive shaft from the tractor that tows it. The machine may be provided with two sets of rubber beater arms to follow up the topping device. The beet can then be cut off at a somewhat higher level above ground (see Fig. 5); the short leaves that are then still left standing on the beet top are removed by the beater arms, which revolve in opposite directions.

Fig. 1

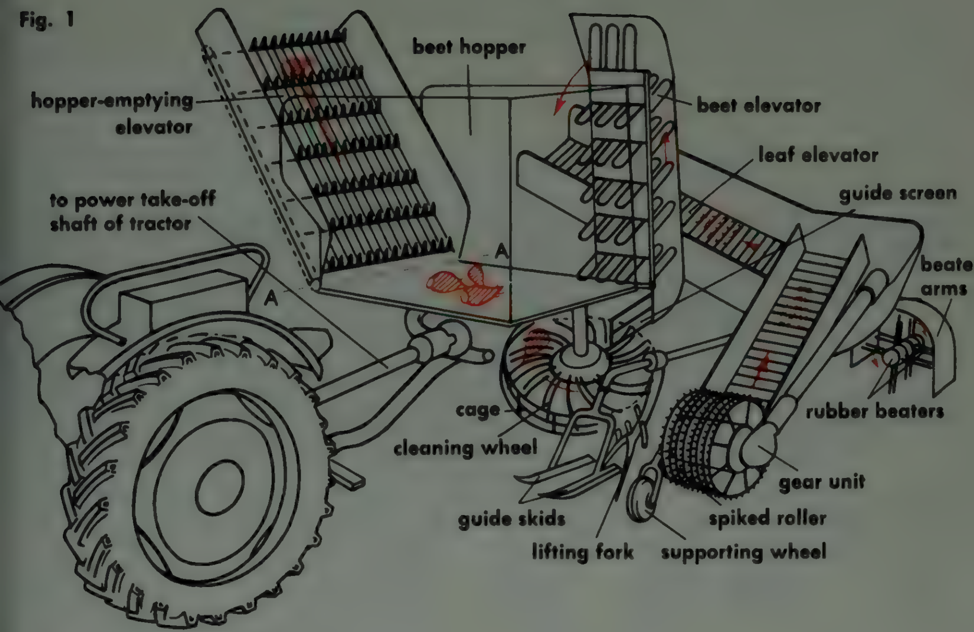
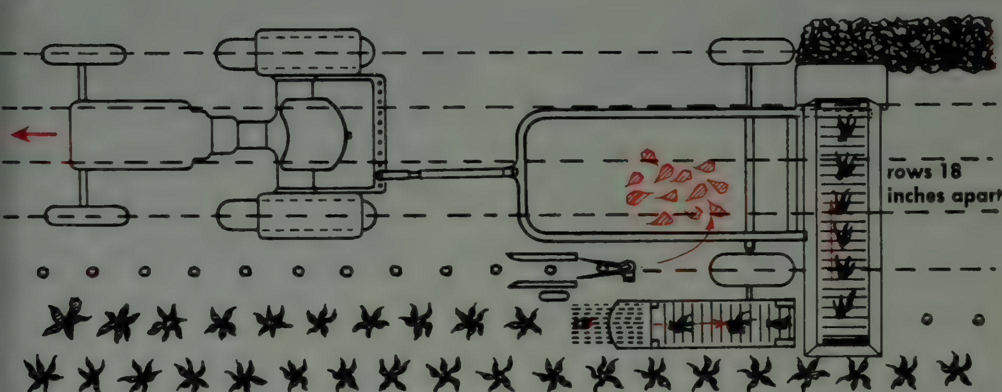


Fig. 2



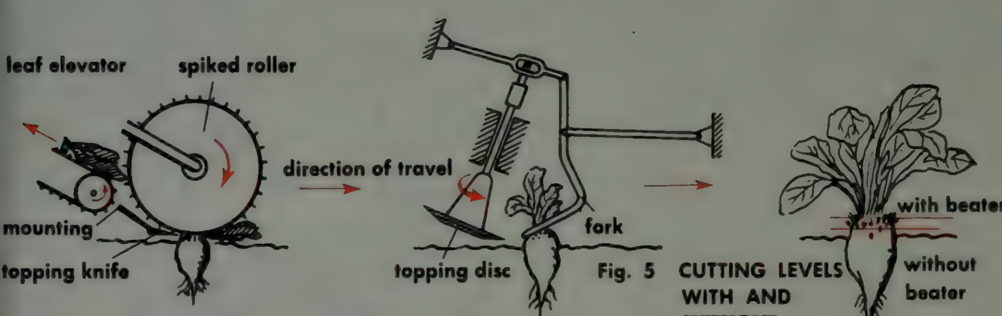


Fig. 3 and Fig. 4 TOPPING DEVICES

Fig. 5 CUTTING LEVELS WITH AND WITHOUT BEATER ARMS

Lettuce picker. As many as 12,000 heads of fresh lettuce can be harvested in one hour by this mammoth portable factory

Photo USIS


BUD

The pick-up baler is used to bale hay (or straw) directly from the windrow in the field. The hay is picked up by prongs and passed to the auger (feed screw) which pushes it to the feed prongs. These deliver the hay to the bale chamber on each stroke of the compressing plunger. The latter compresses the hay into compact layers; protruding stalks which are liable to cause jamming are cut off by knives. A number of layers are tied to form bales (whose length and density can be adjusted) (Fig. 1).

The baler is driven by an internal combustion engine mounted on the baler itself or, alternatively, it is driven by a power-take-off shaft from the tractor that tows it (cf. page 142). The flywheel equalises the power thrusts of the compressing plunger and carries the latter past its "dead-centre" point (i.e., when it is rammed fully home). It thus helps to achieve smoother running. As a safeguard against damage due to the possible presence of stones or pieces of wood or iron in the hay, the flywheel is driven by a shearing pin mounted on the shaft. In the event of overloading, the pin shears through, allowing the flywheel to rotate freely, without transmitting any of its stored-up energy to the plunger. When the obstruction has been removed and the shearing pin replaced by a new one, the machine is ready for operation again. Additional protective devices against overloading are provided in the form of friction clutches which develop slip and thereby shed part of the load before fracture of working parts of the machine can occur.

The bales of hay are automatically tied with wire or twine. The main functioning parts of the twine tying mechanism are illustrated in Figs. 2a–c. The successive stages of the tying operation are shown in Fig. 3. To start with, the twine is threaded through the needle (not shown) and is inserted into the mechanism as indicated in Fig. 3a. While the bale is being compressed and thrust forward by the plunger, the twine is looped round it. When the bale has reached the requisite length, the tying mechanism comes into operation (Fig. 3b). As the plunger is withdrawn, the needle raises the twine, so that the latter encircles the bale, and inserts both ends of the loop into the discs of the twine-holder, which rotate and thus securely hold the twine. At the same time, the knotter begins to turn (Fig. 3c) and, in so doing, opens its jaws so that both ends of the twine are gripped between them (Fig. 3d). The twine-holder discs meanwhile continue their rotation so that the twine attached to the needle is held in the second notch (Fig. 3e). The knotter has now completed its turning motion and its jaws have closed again. The knife swivels round and cuts the twine between the knotter and the twine-holder discs. A continuation of this swivelling movement slips the twine off the ends of the knotter jaws (Fig. 3f), the two cut-off free ends being passed through the loop. The knot has thus been formed, and the knife lever returns to its initial position. Meanwhile the needle descends, thus clearing the way for the next plunger stroke.

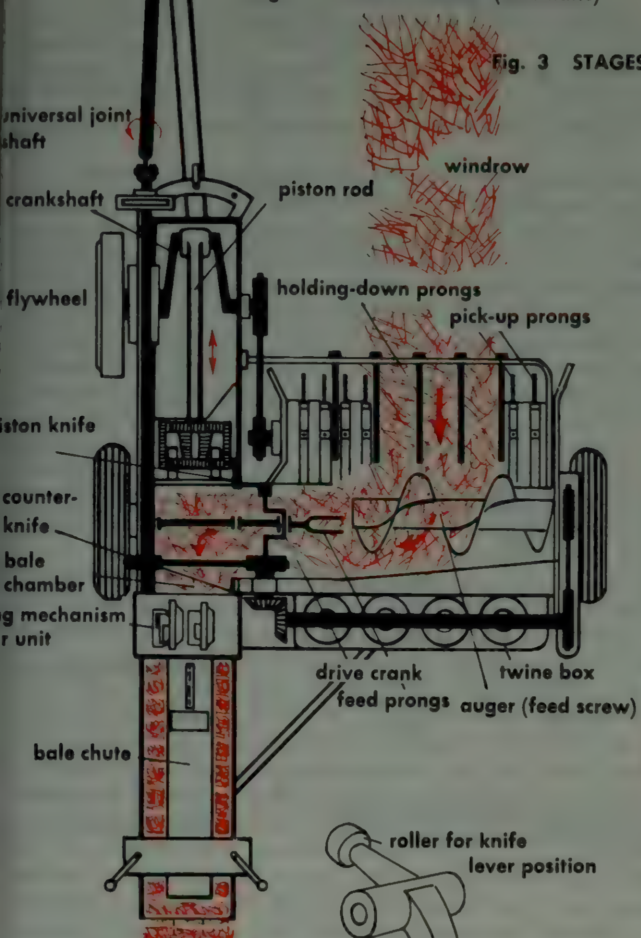


Fig. 1 PICK-UP BALER (*schematic*)

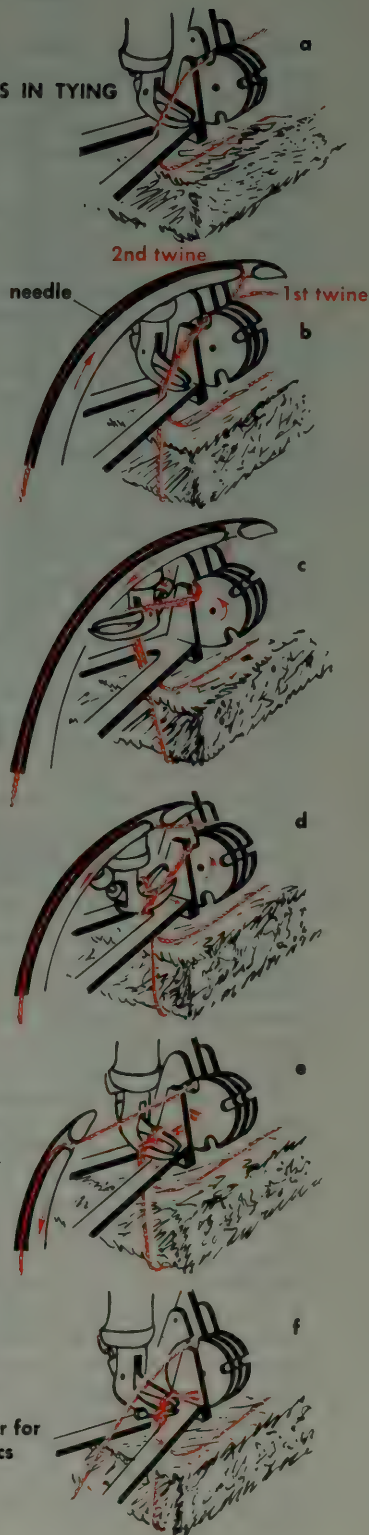


Fig. 3 STAGES IN TYING

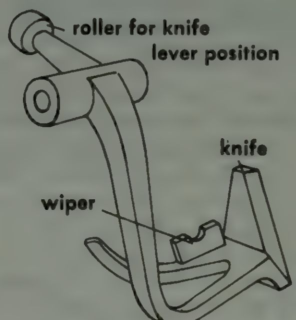


Fig. 2a KNIFE LEVER

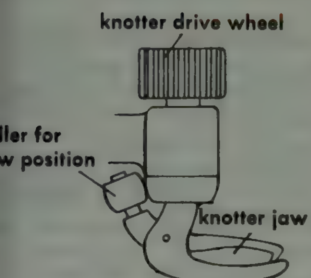


Fig. 2b KNOTTER

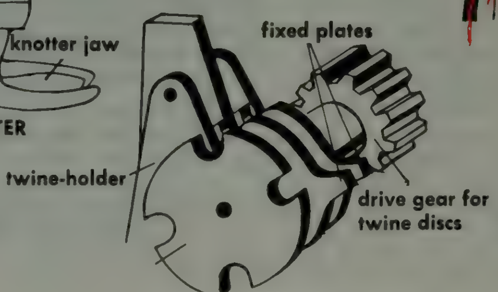


Fig. 2c TWINE-HOLDER

X-RAY APPARATUS AND RADIOLOGY

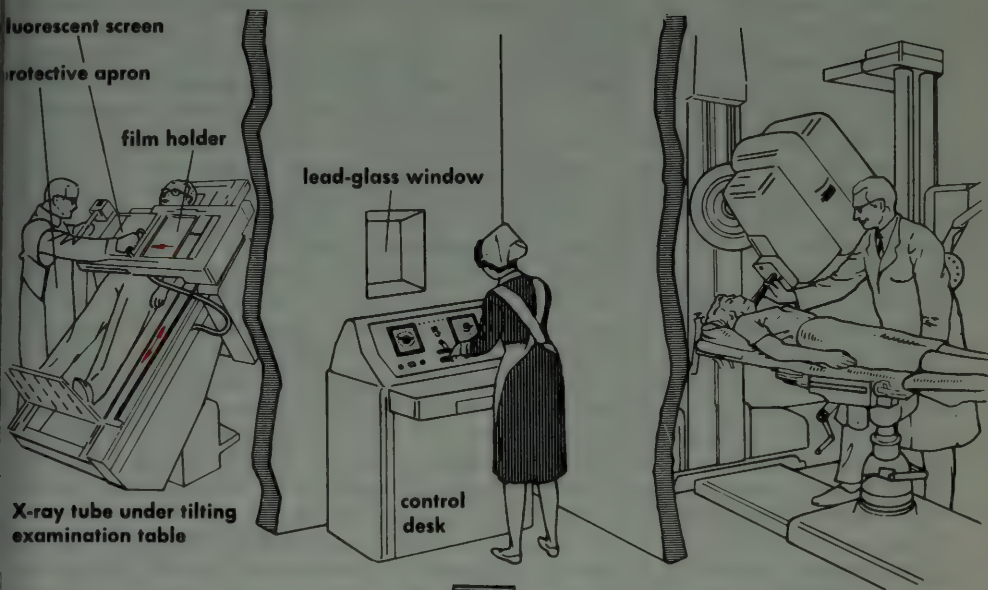
Radiology is the branch of medicine that deals with the use of X-rays in the diagnosis of disease and the use of X-rays, gamma rays and other forms of radiation in the treatment of disease. Thus a major distinction is to be made between diagnostic radiology and therapeutic radiology. Depending on the purposes for which they are intended, a number of different types of X-ray apparatus and X-ray tubes are available.

X-rays are short-wave electromagnetic vibrations which can penetrate solid matter (p.124, vol.I). They are produced when, in a vacuum, electrons are released, accelerated and then abruptly retarded. This takes place in the X-ray tube. To release electrons, the tungsten filament in the tube is heated to incandescence (white heat) by passing an electric current through it. The electrons are accelerated by a high voltage (ranging from about ten thousand to some hundreds of thousands of volts) between the anode (positive) and the cathode (negative) and impinge on the anode, whereby they are abruptly slowed down. In modern X-ray tubes the anode—usually referred to as the "target"—is often of the rotating disc type, so that the electron beam is constantly striking a different point of the anode perimeter. The X-ray tube itself is made of glass, but is enclosed in a protective casing, which is filled with oil to absorb the heat produced. The high voltage for operating the tube is supplied by a transformer (p.110, vol.I). The alternating current is rectified by means of rectifier tubes (or "valves") (Fig. 4) or, in the most up-to-date equipment, by means of barrier-layer rectifiers (see p.96, vol.I).

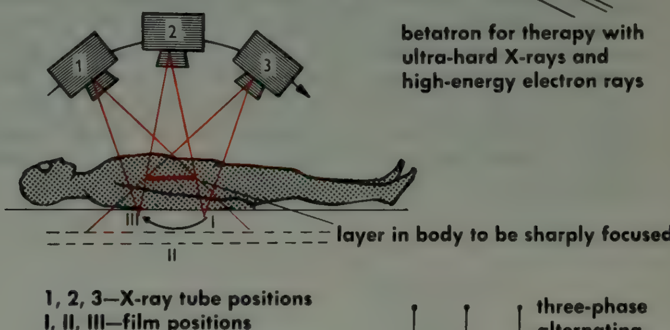
Bones and internal organs into which a suitable contrast medium has been introduced appear most distinctly in X-ray photographs. Various arrangements of the X-ray apparatus and other equipment are employed for the diagnostic examination or therapeutic treatment of various organs. For example, for X-raying the stomach, intestines, heart and lungs an installation as illustrated in Fig. 1 may be used. For X-ray photographs of small appendages, such as teeth, fingers and toes, a small spherical X-ray apparatus, in which the transformer is mounted integral with the tube, may be used. In modern radiology the exposures can be timed by means of multiplier phototubes (p.140, vol.I) (phototiming). X-ray films consist of an acetate cellulose base coated on both sides with an emulsion of silver halide and gelatin. In addition, so-called intensifying screens are used, whereby the contrast of the image can be enhanced. The clinical usefulness of the X-ray examination of certain internal organs (e.g., stomach and intestines) depends on the use of a contrast medium—usually barium sulphate—which is introduced into the organs by various means and makes them show up distinctly in the photograph. Examination by direct visual inspection of the "live" X-ray image on a fluorescent screen, as distinct from photography, is called fluoroscopy. It is not so widely used as photography, as the photographic image shows a greater amount of detail.

For therapeutic purposes—e.g., the treatment of tumours, etc.—the X-rays employed are in some cases generated at much higher voltages (over 4 million volts). Also the rays emitted by radium and artificial radioisotopes, as well as electrons, neutrons and other high-speed particles (for instance, produced by a betatron, Fig. 1), are used in radiotherapy. The tissues of the body are unaffected by radiation that merely passes through them. On the other hand, if a particular tissue absorbs radiation, the cells of that tissue are injured. In the treatment of cancer, for example, the object is to cause maximum damage to the cancer cells with minimum damage to the adjacent normal tissue. The X-ray tube may be moved about, so that, while its radiation remains focused on the deep-seated tumour, it does not constantly pass through the same intermediate tissues. In diagnostic radiology the technique known as tomography, consisting in taking "layerwise" X-ray photographs, is sometimes employed. This is based on the use of rays varying in penetrating power and thus photographing the interior of the body at different depths, thus enabling the exact location of an infection (e.g., in the lungs) to be ascertained. For this purpose the X-ray tube and the film may be moved in opposite directions in curved paths (Fig. 2).

Fig. 1 EQUIPMENT FOR DIAGNOSTIC AND THERAPEUTIC RADIOLOGY



ig. 2 TOMOGRAPHY



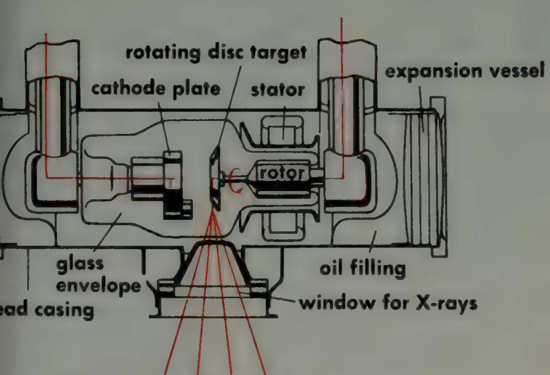


Fig. 3 X-RAY TUBE IN OIL-FILLED PROTECTIVE CASING

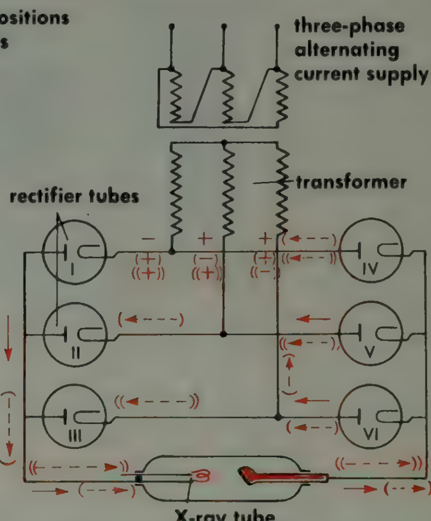
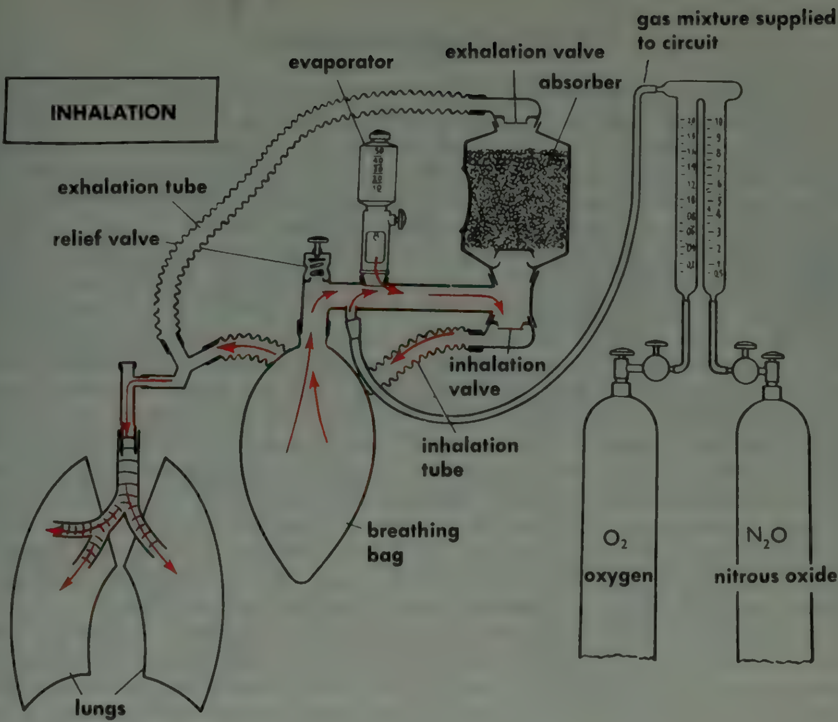


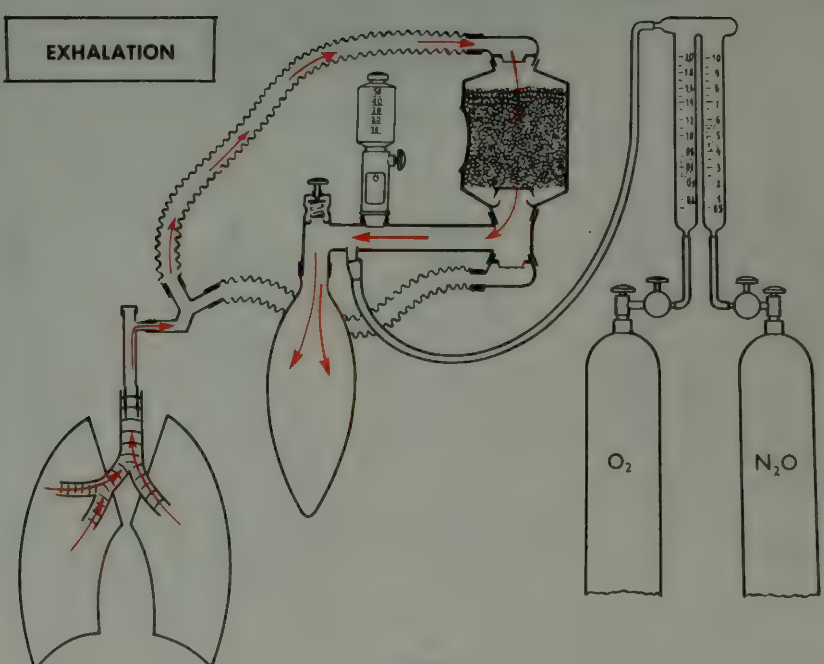
Fig. 4 CIRCUIT DIAGRAM OF A SIX-TUBE INSTALLATION

ANAESTHETIC APPARATUS

Anaesthesia means loss of feeling or sensation, so that no pain is felt. In surgery this result is obtained by using an anaesthetic. A distinction is to be made between general anaesthesia (total unconsciousness) and local anaesthesia (only one area of the body is deprived of sensation).

General anaesthesia can be produced in various ways, including intravenous injection with thiopental sodium or other agents. The older and still most widely employed method, however, is by inhalation of a gaseous or volatile anaesthetic. Early anaesthetics were ether, nitrous oxide, and chloroform. At the present time a range of other agents are available. In modern surgery, especially for major operations, a combination of two or more anaesthetic agents may be employed, the gaseous or volatile anaesthetic being administered by means of a special apparatus, which enables the various agents to be accurately proportioned and controlled, so as to minimise the risk of overdosing. A typical anaesthetic apparatus is shown in the accompanying illustration. The underlying principle is that the patient's breath is circulated through the apparatus, in a closed circuit, the gas flow rate being controlled by means of valves and flow meters. The advantage of the closed circuit is that loss of body heat and moisture is prevented. Besides, a considerable economy in the amount of anaesthetic used is effected. Also included in the circuit are a breathing bag (which shows the breathing movements and their frequency and which can be squeezed in order to intensify the inhalation), an inlet attachment for supplying fresh air, an evaporator for volatile anaesthetic agents (e.g., diethyl ether) should these be used, and a cartridge containing an absorbent for the carbon dioxide contained in the exhaled air. This air may be recycled through the breathing circuit or may, in other varieties of anaesthetic apparatus, be discharged from the apparatus. The anaesthetic is administered to the patient either through a face mask or through a tube introduced into the trachea (windpipe), the latter method now being considered safer and more effective. In the apparatus illustrated, the anaesthetic is nitrous oxide gas used in conjuction with the vapour of a volatile anaesthetic. Before having this mixture administered to him by inhalation the patient is usually given a preliminary anaesthetic by intravenous administration, i.e., in the form of an injection.


INHALATION
gas mixture supplied to circuit
exhalation valve
evaporator
absorber
exhalation tube
relief valve
inhalation valve
inhalation tube
breathing bag
lungs
O₂
oxygen
N₂O
nitrous oxide


EXHALATION
O₂
N₂O

Every muscle can perform only one movement, namely, the shortening of its fibres by contraction. This also applies to the heart muscle. Each action of a muscle is associated with electric currents which change in the course of the contraction and which, after passing through the various tissues, reach the surface of the body. There they can be picked up by electrodes and be recorded with the aid of a suitable apparatus, the electrocardiograph. The record thus obtained on a chart is called an electrocardiogram. In its original form the instrument was based on the principle of the string galvanometer, which was invented in 1903 by Einthoven, a Dutch physiologist (Fig. 1). A silver-plated quartz filament is brought into a magnet field. The electric currents produced by the contraction of the heart muscle and picked up by the electrodes applied to the patient's body are passed through this filament, which undergoes a deflection whose direction and magnitude will depend on the direction and strength of the current. This movement of the filament is projected, by means of an optical system, as a spot of light on to a moving paper strip chart which is coated with a light-sensitive compound. When no current is flowing through the filament, a straight line is traced on the chart. Currents associated with the muscular action of the heart cause the spot of light to oscillate and thus trace a typical curve on the light-sensitive chart. Any irregularities in the functioning of the heart appear as corresponding irregularities in the curve which enable the heart specialist to diagnose the disease or other cause of these deviations from the normal pattern.

The string galvanometer instrument has now been superseded by the electronic cardiograph which operated with amplifier tubes (Fig. 2). The currents from the body electrodes enter the amplifier at E_1 and E_2. The greatly amplified currents are fed to a mirror galvanometer. The movements of the mirror cause a reflected spot of light to oscillate on the light-sensitive chart (film) and thus trace the required curve.

The electrodes are affixed to the human body at certain definite points (Fig. 3): left arm and right arm (1), left leg and right arm (2), left leg and left arm (3). These were the points originally selected by Einthoven. Nowadays other points of attachment to the limbs and also to the wall of the chest are used. Modern cardiograph records do not record these various curves separately, in successive operations, but record them simultaneously. The normal electrocardiogram presents the appearance shown in Fig. 4. The portion of the curve between *P* and *Q* corresponds to the contraction of the auricles, while the portion between *Q* and *T* corresponds to the contraction of the ventricles of the heart.

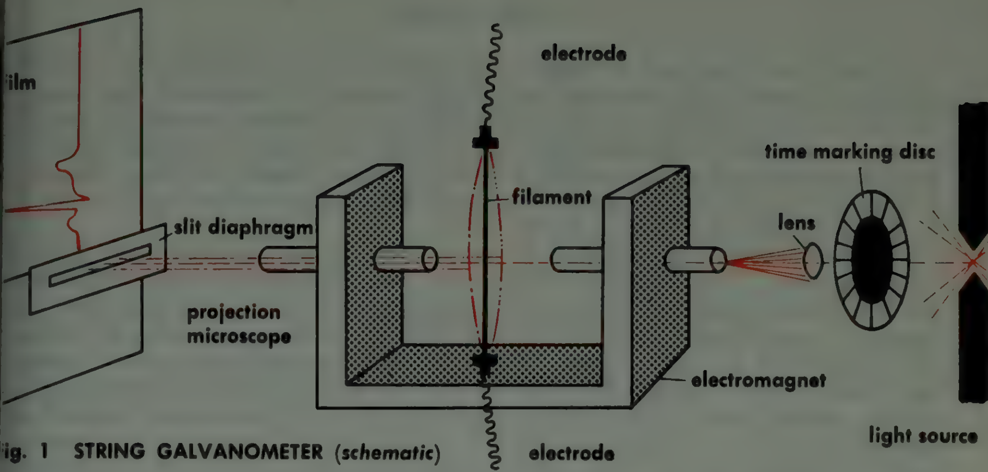


Fig. 1 STRING GALVANOMETER (*schematic*)

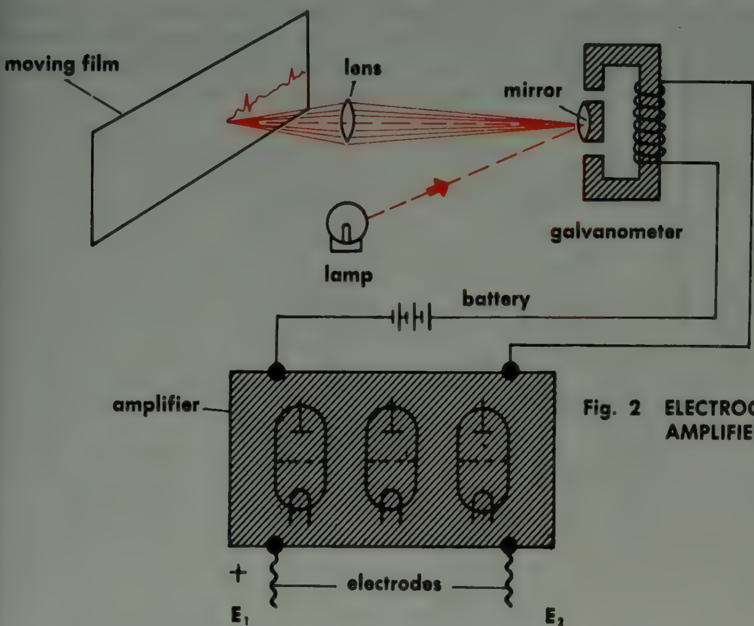


Fig. 2 ELECTROCARDIOGRAPH WITH AMPLIFIER SYSTEM

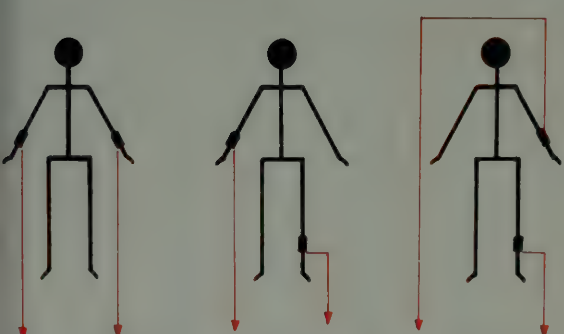


Fig. 3 POINTS OF ATTACHMENT FOR ELECTRODES (*according to Einthoven*)

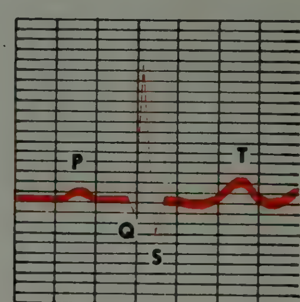
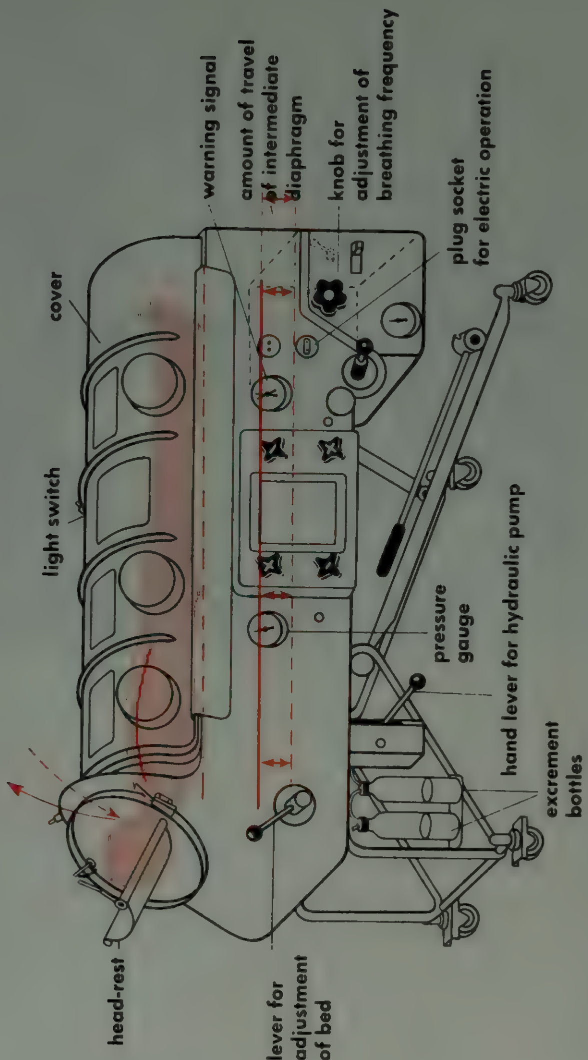


Fig. 4 NORMAL ELECTROCARDIOGRAM

In cases where paralysis or other causes impair normal breathing, the machine popularly known as the iron lung can help the patient to breathe and thus remain alive indefinitely or until his recovery. In the iron lung an excess pressure (in relation to the atmospheric air pressure) alternates rhythmically with a reduced pressure. When the pressure surrounding the patient's body is reduced, his chest expands, so that air streams into his lungs. Then when the pressure is increased, the air is automatically expelled from the lungs.

In the machine illustrated (Type E 52 manufactured by the German firm of Dräger) the patient lies on a foam-rubber-cushioned bed, with his head protruding out of the end of the machine and reclining on an adjustable head-rest. At the other end is a likewise adjustable foot-rest. In the chamber under the bed is a movable intermediate diaphragm which is moved up and down by the drive mechanism (powered by electricity) and thus performs a bellows function. As the chamber is airtight, this alternate increase and decrease of its volume produces the reduced pressure and excess that enables the patient to breathe in and out. In the event of a power cut, a warning signal is sounded. The iron lung can then be operated by hand. The rate of rise and fall of this intermediate diaphragm determines the breathing frequency. For washing or examining the patient the cover of the machine can be swung open. The patient's head is then temporarily enclosed in a plastic dome in which the air pressure is alternately raised and lowered, so that he can continue to breathe.

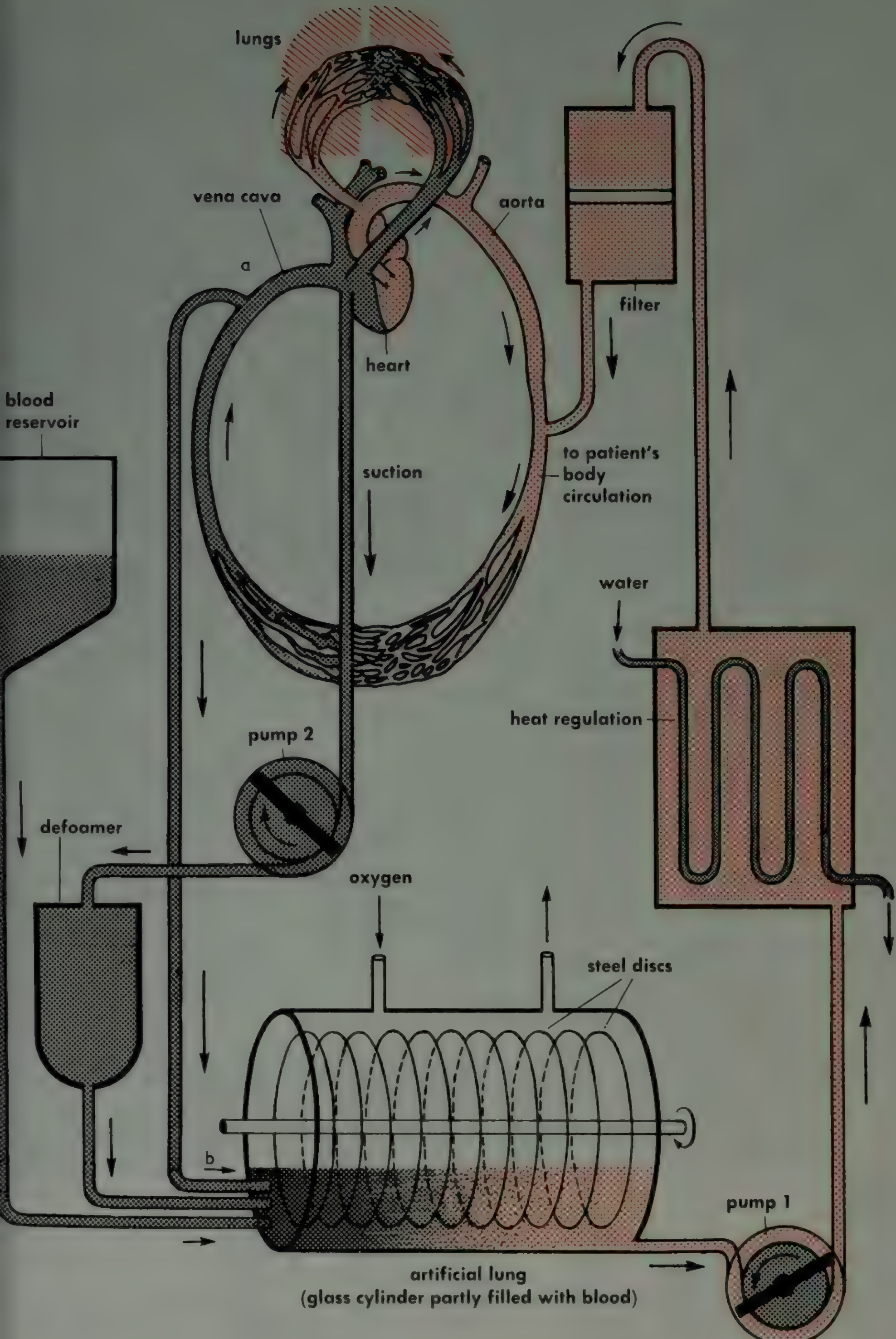

cover
light switch
head-rest
lever for adjustment of bed
warning signal
amount of travel of intermediate diaphragm
knob for adjustment of breathing frequency
plug socket for electric operation
pressure gauge
hand lever for hydraulic pump
excrement bottles

"Iron lung" respirator
Photo Hôpital Cantonal, Lausanne, Switzerland

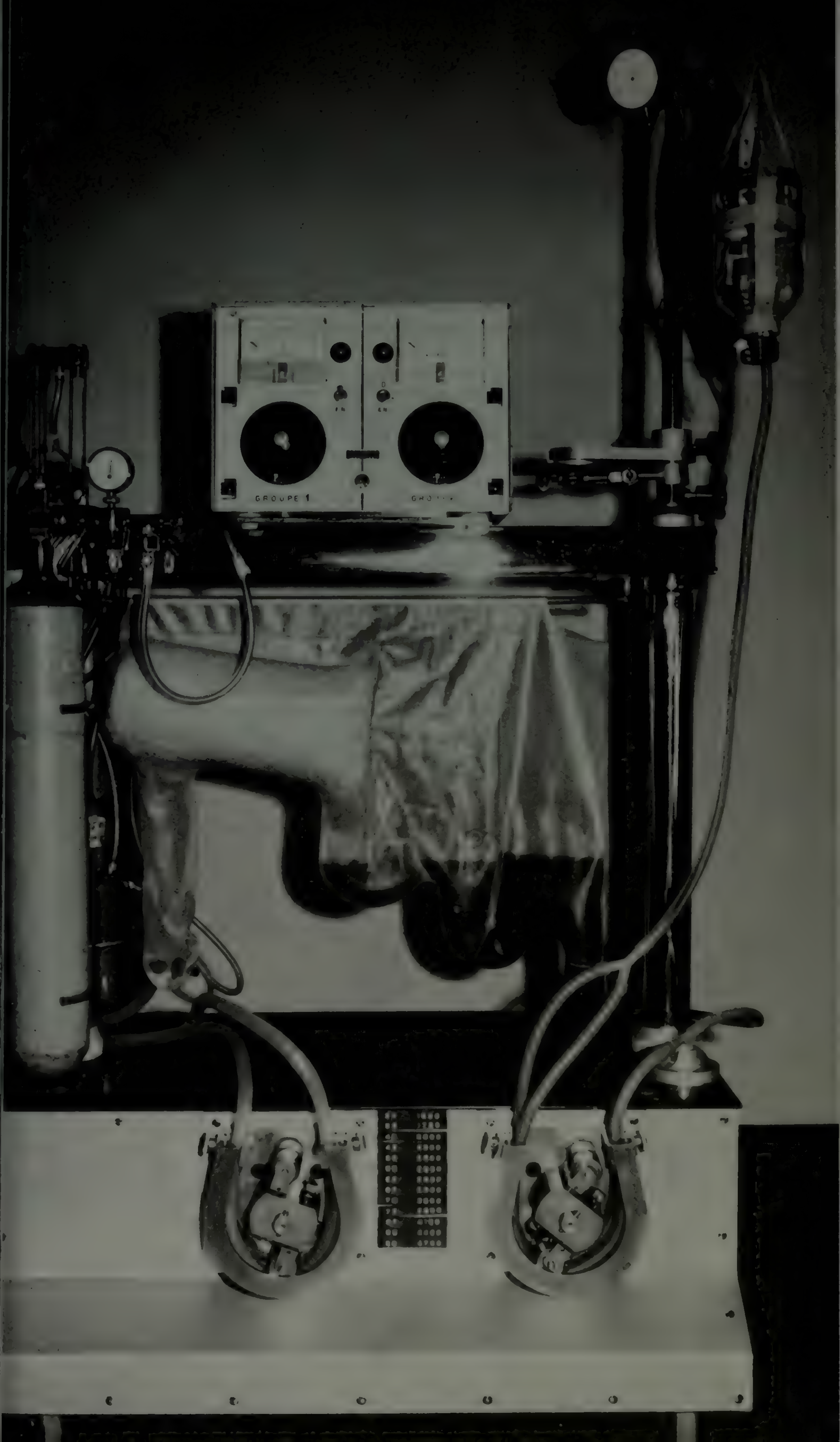

ENGSTRÖM RESPIRATOR modell 200

Surgery of the heart usually involves opening the heart muscle. For operations of short duration it is possible to apply hypothermy (deep cooling) and thus temporarily stopping the blood circulation altogether. However, for major operations it is necessary to maintain the circulation, and this is achieved by means of the heart-lung machine. This has the twofold function of keeping the replacement blood in circulation by means of a pumping system and of enriching with fresh oxygen the blood of low oxygen content coming from the patient's body. The system described here is the one used at Heidelberg University Hospital and elsewhere.

The venous blood, before it enters the right auricle of the heart, is diverted out of the vena cava and passed into plastic tubes (a). This blood, which has already circulated through the body and consequently has a low oxygen content, is circulated through an artificial lung (b). In a horizontal glass cylinder partly filled with blood a number of steel discs rotate, which thus become wetted with blood. The blood on the surface of these discs forms a thin film of large area, which is exposed to a stream of oxygen in the upper part of the glass cylinder. The red blood cells are thus able to absorb oxygen in much the same way in which they do this in the human lung. The pump 1 now passes the oxygen-saturated blood through a heat controller and a filter and then back to the patient's arterial circulation. Losses of blood occurring in the course of the operation are compensated by a blood reservoir. A second pump (pump 2) extracts venous blood from the heart itself, this being blood reaching the heart through veins other than the vena cava. This blood is defoamed and likewise passed to the artificial lung. Before starting, the machine is filled with three or four litres of blood to which an anti-coagulant has been added so that it cannot congeal. All the internal parts of the apparatus (except the steel discs) are silicone-treated to make them unwettable.


lungs
vena cava
aorta
a
heart
filter
blood
reservoir
suction
to patient's
body
circulation
water
heat regulation
pump 2
defoamer
oxygen
steel discs
b
pump 1
artificial lung
(glass cylinder partly filled with blood)

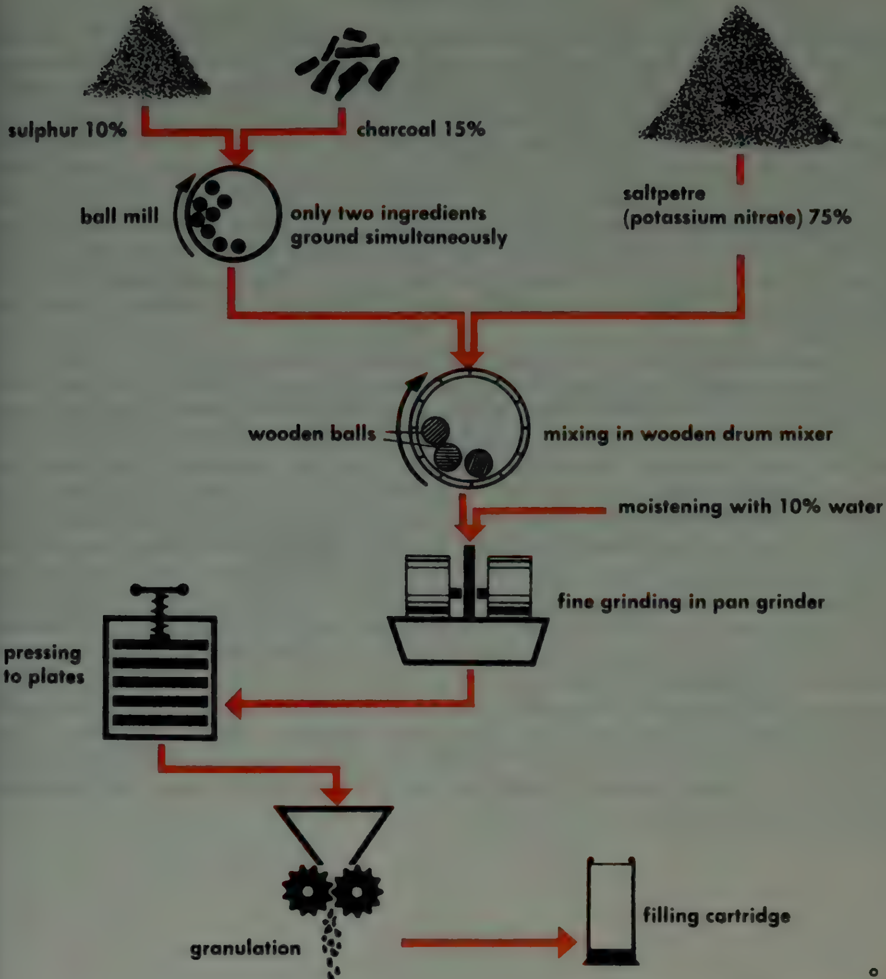
Heart-lung machine
Photo Hôpital Cantonal, Lausanne, Switzerland

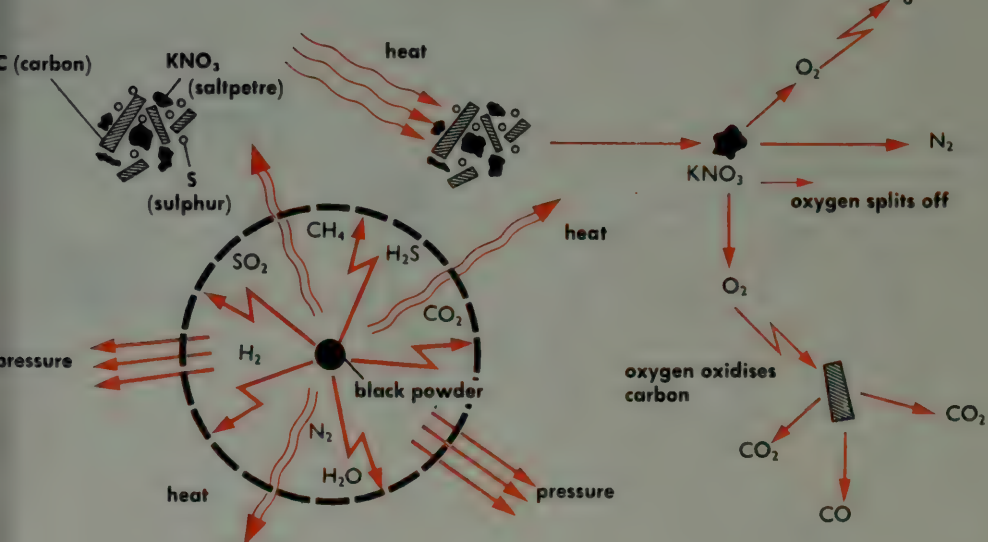

GROUPE 1

Gunpowder is an explosive which has a relatively low detonation velocity. Its action is propellant rather than shattering. The general term "propellant" is applied to any explosive which can suitably be used for the propelling of projectiles fired from guns or of rockets. These explosives consist either of intimate mixtures of substances which react with one another and release a considerable amount of energy while doing this, or of chemical compounds which release energy on decomposition. In both cases, however, the reaction does not take place at ordinary temperatures: ignition of the explosive is necessary, i.e., the reaction has to be initiated by supplying energy at one point. Once the reaction has thus been started, the energy that is released initiates the reaction at adjacent points. This "chain reaction" can spread throughout the whole quantity of explosive in a small fraction of a second.

The significant requirement for a suitable propellant is that a considerable increase in volume shall occur during the course of the reaction. The reaction products are for the most part gases which occupy a much larger volume than the solid explosive. If the reaction takes place in a confined space (as, for example, in a cartridge fired in a rifle), a very high pressure therefore develops, which drives the projectile out of the barrel. The oldest known propellant is, of course, ordinary black powder (gunpowder). When this explodes, the reaction forms about 45% gases (nitrogen, carbon monoxide, carbon dioxide) and 55% vaporised salts. A pound of gunpowder will thus produce about 5 cubic feet of gas and about 300 kilocalories (1200 B.T.U.) of heat. The bottom left-hand diagram shows how much gas (the large circle) is evolved from the explosion of a small quantity of powder (represented by the black dot).

Gunpowder is a mixture of granular ingredients, namely, sulphur, potassium nitrate (saltpetre) and charcoal. The ingredients are ground separately or two by two, then pressed into cakes in a moistened condition, and finally reduced to granules which are used for the filling of cartridges, fireworks, fuses (for blasting, etc.). In modern fire-arms gunpowder has largely been superseded by nitrocellulose-based and other smokeless explosives. To make these suitable for use as propellants, the nitrates are dissolved in solvents and formed into various shapes (threads, tubes, small plates, etc.) which assist rapid and efficient burning of the explosive.


sulphur 10%
charcoal 15%
ball mill
only two ingredients
ground simultaneously
saltpetre
(potassium nitrate) 75%
wooden balls
mixing in wooden drum mixer
moistening with 10% water
fine grinding in pan grinder
pressing
to plates
granulation
filling cartridge

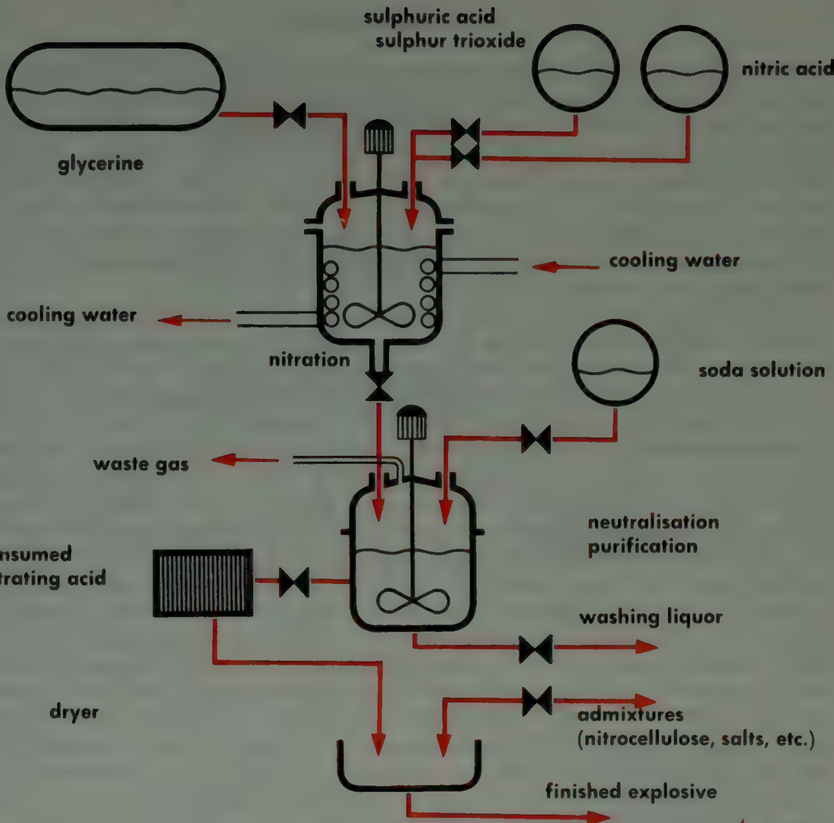

C (carbon)
KNO_3
(saltpetre)
S
(sulphur)
heat
SO_2
O_2
N_2
KNO_3
oxygen splits off
CH_4
H_2S
SO_2
CO_2
H_2
black powder
N_2
H_2O
pressure
heat
heat
pressure
O_2
oxygen oxidises
carbon
CO_2
CO_2
CO

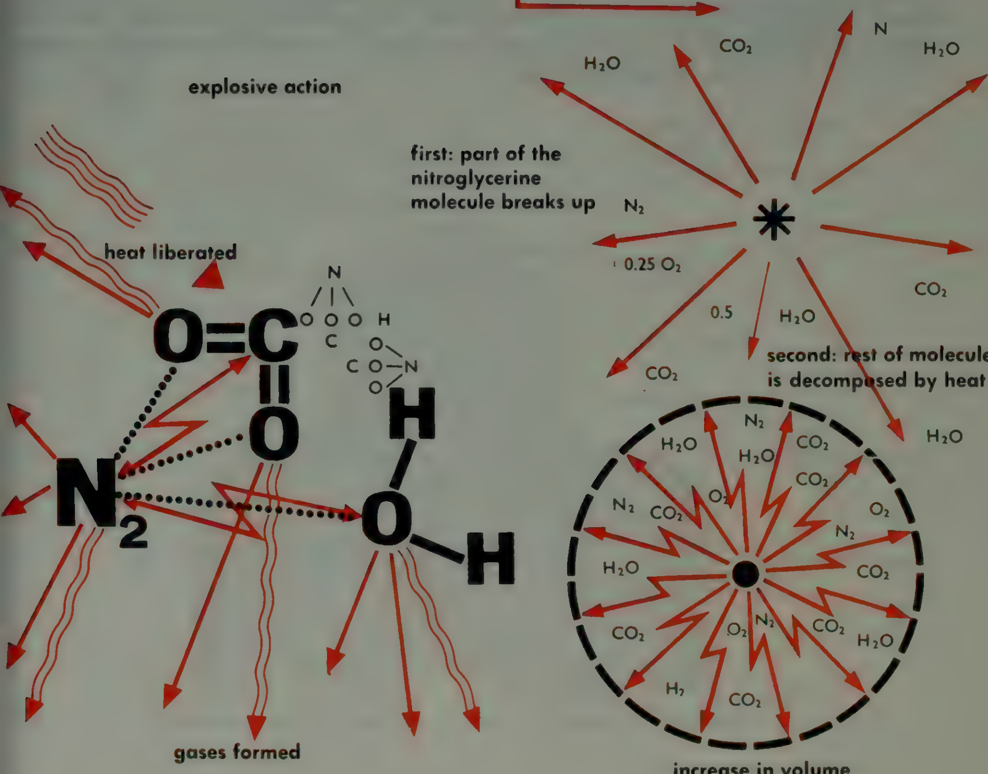
The explosives more particularly considered here are of the "brisant" or "shattering" type which explode on heating as a result of a blow, or friction, or ignition and which then develop a very rapid and shattering action. These explosives are used more particularly for blasting in quarrying, mining, tunnelling, demolition work, etc., i.e., in cases where large masses of solid material have to be broken up quickly and cheaply. In general they are too violent and therefore unsuitable as propellants.

By varying the composition of the explosive, its "brisance" (shattering power) can be adjusted to suit the particular purpose for which it is to be used. The various explosives differ in the amount of gas that is produced per pound of explosive, in the amount of heat liberated per pound, in the detonation velocity, in the pressure that the hot gases exert upon their immediate surroundings, and in the shattering power developed.

For instance, one pound of blasting gelatine (nitroglycerine with about 8% nitrocellulose added) exploding at a temperature of over 4700° C produces about 680 kilocalories (2700 B.T.U.) of heat, over 9 cubic feet of gas is formed, and a pressure of about 13,000 atm (190,000 lb./in.^{2}) is developed as the explosion takes place in an enclosed space. Although the amounts of heat and gas liberated in the explosion are not very much than twice those produced in the explosion of one pound of gunpowder, which is a much slower explosive (see page 170), the shattering power of blasting gelatine (and similar explosives) is very much greater than that of gunpowder, this being due mainly to the greater suddenness of the explosion reaction.

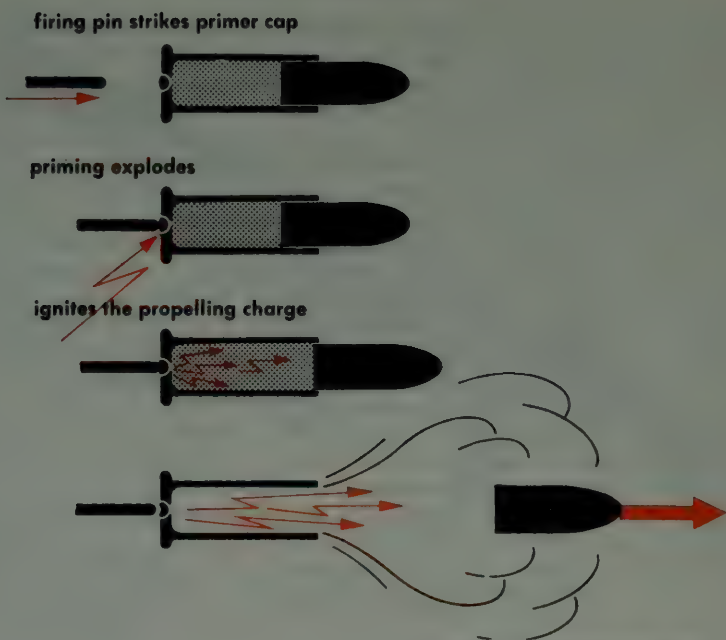
Dynamite as originally invented by Nobel consisted of 75% nitroglycerine (glycerol trinitrate, a yellow liquid) compounded with 25% kieselguhr (a diatomaceous earth consisting of the tiny shells of microscopic marine creatures). The term "dynamite" is nowadays used in a wider sense to comprise various explosives consisting of nitroglycol, ammonium nitrate, saltpetre, aromatic nitro compounds, and wood meal. By adding up to about 40% common salt to dynamites, their shattering power is so reduced that they become safe for use in mining because they can then be exploded without attendant risk of igniting any mine gas (methane) or coal dust that may be present in the air.

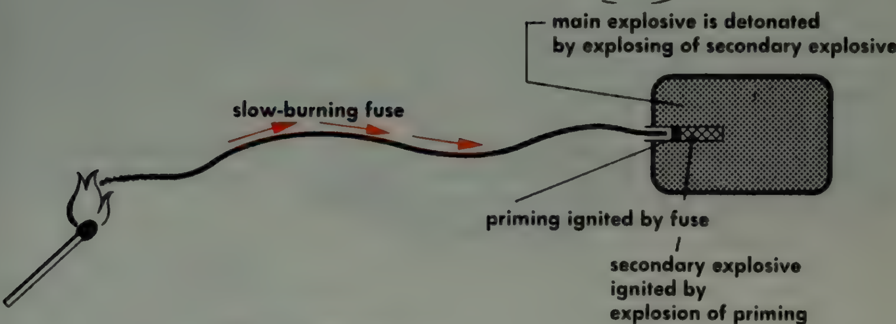

sulphuric acid
sulphur trioxide
nitric acid
glycerine
cooling water
cooling water
nitration
soda solution
waste gas
neutralisation
purification
consumed
nitrating acid
washing liquor
dryer
admixtures
(nitrocellulose, salts, etc.)
finished explosive

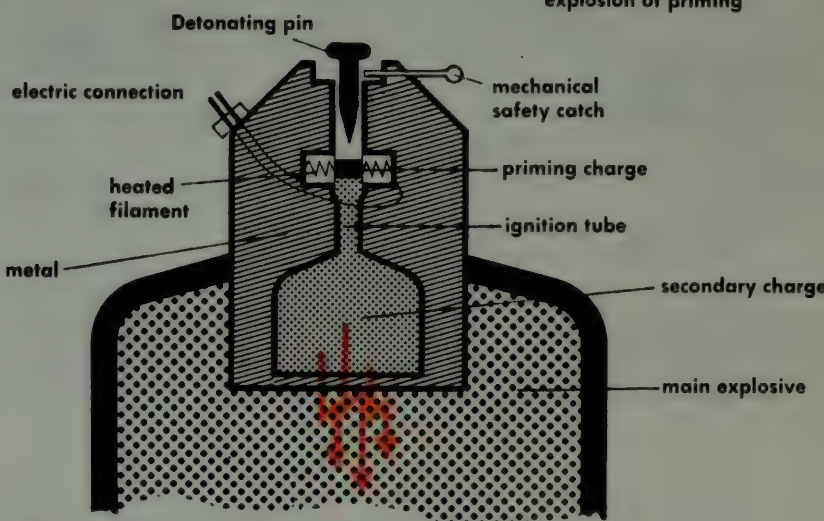

explosive action
first: part of the
nitroglycerine
molecule breaks up
heat liberated
second: rest of molecule
is decomposed by heat
gases formed
increase in volume

To be safe and efficient in use, explosives should explode only when they are required to do so. Under ordinary conditions of handling they should be stable, insensitive to pressure, impact and heat, and even (in some cases) withstand the effect of sparks or a small flame without exploding. On the other hand, they should be highly effective on being exploded. Alfred Nobel, in 1867, found that a "safe" (insensitive) explosive can be efficiently exploded by detonating it—initiating its explosion—with a small quantity of a highly sensitive, and in itself therefore dangerous, explosive. Such explosives (called "priming explosives" or "primers") explode easily on being subjected to a blow, heating or friction.

There are various types of detonating devices depending on the purpose for which they are employed: for blasting, for military use (in cartridges, artillery ammunition, bombs, etc.). Detonation can be effected by heat (fire), by an electric spark, by impact, etc. An ordinary small-arms cartridge is usually of the centre-fire type, i.e., the metal cartridge case containing the propellant is provided at the centre of its base with primer cap containing a small quantity of priming which is exploded by a blow from the hammer or firing pin. For ordinary blasting as used in quarrying, mining, tunnelling, etc. the main explosive is detonated by means of a detonator (or blasting cap). This consists of a small metal tube containing a small quantity of priming explosive, e.g., mercury fulminate or lead azide, together with a larger quantity of a secondary explosive such as trinitrotoluene (T.N.T.). The detonator may be designed for electric detonation, for which purpose it is provided with a filament in the priming. Alternatively, a detonator may be ignited by a slow-burning fuse (generally called "safety fuse") inserted into one end of the detonator or by so-called detonating cord or fuse, which burns at a very much higher speed and is particularly useful in cases where two or more charges have to be detonated more or less simultaneously or in accurately timed rapid succession. The cord consists of a plastic sheath enclosing a core of fast-burning explosive (the burning rate is about 20,000 ft./sec., whereas the ordinary safety fuse burns at only about 2 ft./sec.) and must itself be ignited by a detonator. Modern artillery shells are sometimes fitted with relatively elaborate fuse and detonation equipment designed to ignite the charge by impact and/or after a certain predetermined length of time (time fuse).







principle of an electrically or mechanically operated detonator

An automatic weapon can fire a number of rounds in quick succession, all the functions of firing and reloading being performed by the weapon itself: firing the cartridge, withdrawing the bolt, ejecting the spent cartridge case, cocking the hammer, forcing the bolt forward, and inserting a fresh cartridge into the chamber ready to fire. The energy for performing these functions is provided by the pressure of the gas produced by the firing of the cartridges (gas-operated weapons) or by the recoil of the weapon. Well-known gas-operated rifles are, for example, the Garand rifle, which was widely used by the United States forces in World War II, and the more recent FN rifle, of Belgian origin, which was adopted by the British army in 1954.

The rifle illustrated in Fig. 1, on the other hand, is a recoil-operated weapon—designated as the G3—and designed to fire the NATO standard 7.62 mm calibre cartridge. Its operating principle is shown in Figs. 2 and 3. To start with, the weapon is loaded, cocked and ready for firing. When the trigger is pulled, the cocked hammer is released and strikes the firing pin whereby the cartridge is fired. The gases produced by the explosion propel the bullet out of the barrel. At the same time the gases thrust the cartridge case and, with it, the breech block backwards. The gas pressure is so great that the sealing rollers are forced against the sloped shoulders of the bolt. The latter slides backwards, and the rollers are withdrawn into the breech block until the bolt is disengaged. The rear shoulders of the bolt strike the bolt carrier and thrust it backwards. In its backward movement the extractor of the breech block extracts the cartridge case (which is ejected by the ejector) and cocks the hammer. When the breech block has completed its backward motion, it is returned forward by the action of a spring. In the course of this return movement the end face of the breech block strips the top cartridge from the magazine and thrusts it into a chamber ready to fire. The extractor engages with the corresponding groove at the base of the cartridge. The sealing rollers are pushed outwards by the sloped shoulders of the bolt until they engage with the recesses in the breech wall. The firing cycle has thus been completed and the rifle is ready to fire the next round. The rifle can be operated either automatically—i.e., so long as the trigger is kept pulled, the weapon continues to fire until the magazine is empty—or as a semi-automatic (self-loading) rifle, a separate squeeze of the trigger being necessary to fire each shot.

Fig. 1 G 3 AUTOMATIC RIFLE
7.62 mm calibre

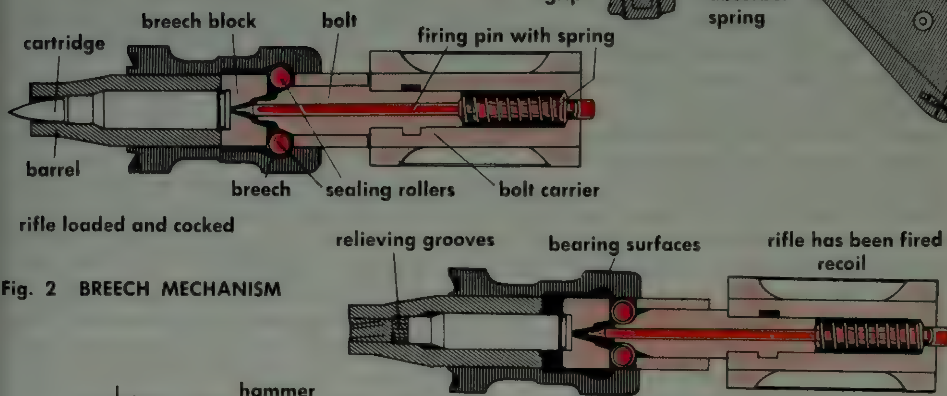
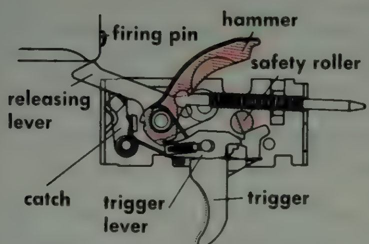
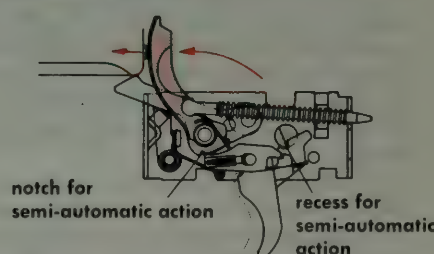


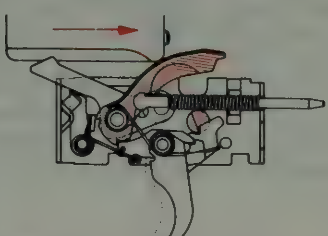
Fig. 2 BREECH MECHANISM



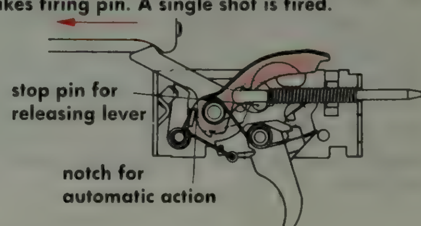
a) Trigger pulled back to initial stop. Hammer ready for semi-automatic fire.



b) Trigger pulled farther back. Hammer strikes firing pin. A single shot is fired.



c) After firing. Recoil. Position of hammer just before engagement of catch. Trigger lever acts only with semi-automatic fire.



d) Safety roller set to "automatic." The trigger has to be pulled back a longer distance. Trigger lever does not act now. Hammer is released by means of releasing lever instead.

Fig. 3 TRIGGER MECHANISM

Radioactivity is the spontaneous disintegration of certain heavy elements (radium, uranium and others) accompanied by the emission of high-energy radiation which consists of three kinds of rays: alpha-rays, beta-rays, and gamma-rays. The ultimate end product of radioactive disintegration is one of the isotopes of lead. The radiation can be split up into its three components by passing it through a magnetic field (Fig. 1). Alpha-rays consist of alpha-particles, which are the nuclei of helium atoms which are positively charged on account of having lost two electrons. Beta-rays consist of beta-particles, which are electrons (negatively charged elementary particles). The positively charged alpha-particles and the negatively charged beta-particles are deflected in opposite directions by the magnetic field. Gamma-rays are an electromagnetic radiation of very short wavelength and high penetrating power (about 10^{-12} cm); they are not deflected by the magnetic field. All three kinds of radiation cause blackening of a photographic plate and ionisation of gases, so that they can be detected by photographic or electric methods.

Radioactive disintegration can be visualised as follows (according to G. Gamow): The nucleus of the radioactive atom is conceived as a kind of pot-like receptacle (Fig. 2) containing alpha-particles, the basic components of the nucleus. Normally these particles can get out of the pot only by flying over the edge. However, alpha-particles are not merely particles of matter; they also possess wave properties which enable them (in limited numbers) to penetrate the wall of the pot. Particles escaping in this way constitute the alpha-radiation. Beta-radiation can be similarly explained, whereas gamma-radiation is associated with changes in energy states inside the atomic nucleus in conjunction with two other kinds of radiation. Fig. 3 shows diagrammatically an apparatus for the detection of alpha-rays which is known as the Geiger-Müller counting tube (or simply "Geiger counter"). It detects the electrically charged alpha-particles by virtue of the ionisation of a gas. The tube is filled with a rarefied gas in which the radioactive radiation initiates a process called impact ionisation whereby a voltage impulse of short duration is produced in the external circuit. This impulse is amplified and fed to the counting device. The high sensitivity of the Geiger counter is due to the geometrical configuration of its electrodes. The thin wire anode (positive pole) is concentrically enclosed by the large cylindrical cathode (negative pole). The probability that the electrons (negative particles) which are released in the impact ionisation process will immediately fly to the anode is slight. The great majority of the electrons rush past the anode and are directed into spiral paths by the powerful electric field of force around the anode. In travelling along these paths the electrons initiate a large number of further impact ionisation processes before they end their flight by finally reaching the anode. The ionisation in the tube gives rise to the impulse by causing a momentary discharge of electricity in the tube.

In addition to the natural radioactivity of radium and uranium there is artificially induced radioactivity, by which is understood a state of radioactivity induced in normally non-radioactive elements by means of artificial atomic transmutation. Such elements are now obtained as waste products from nuclear reactors (p. 68, vol. I). All radioactive phenomena die away after a certain length of time. The so-called half-life (or half-decay period) has been introduced as a criterion of this. It denotes the period in which the activity of a radioactive substance falls to half its original value. It varies considerably for different substances and ranges from a vast number of years to a tiny fraction of a second. The half-lives of the radioactive fission products of a nuclear explosion are mainly in the range of 10–60 days, but there are also much longer-lasting products such as strontium 90 and caesium 137, which have half-lives of about 30 years.

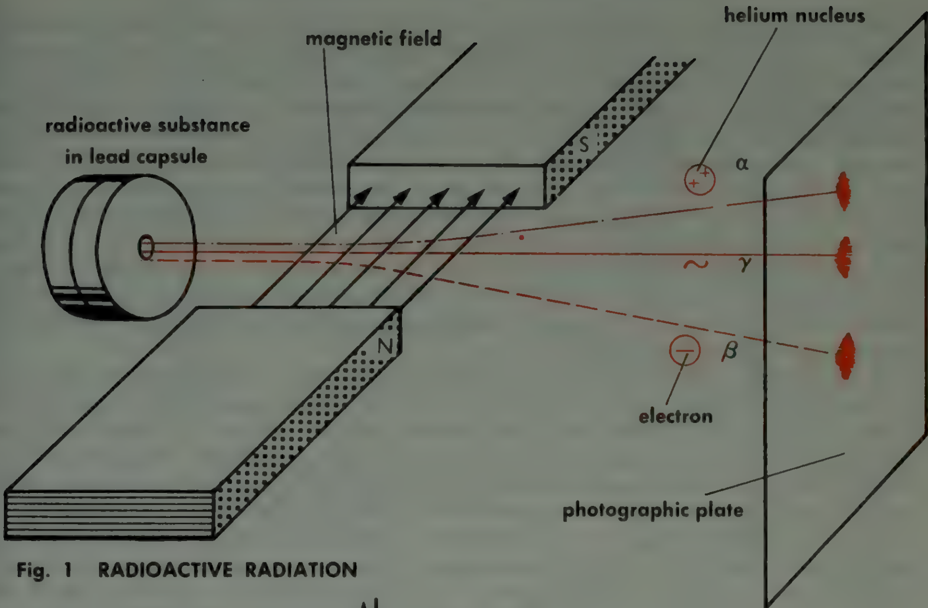


Fig. 1 RADIOACTIVE RADIATION

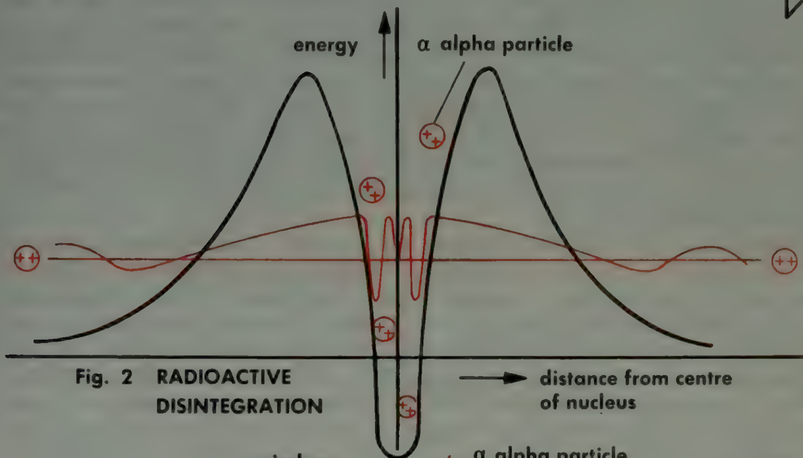


Fig. 2 RADIOACTIVE DISINTEGRATION

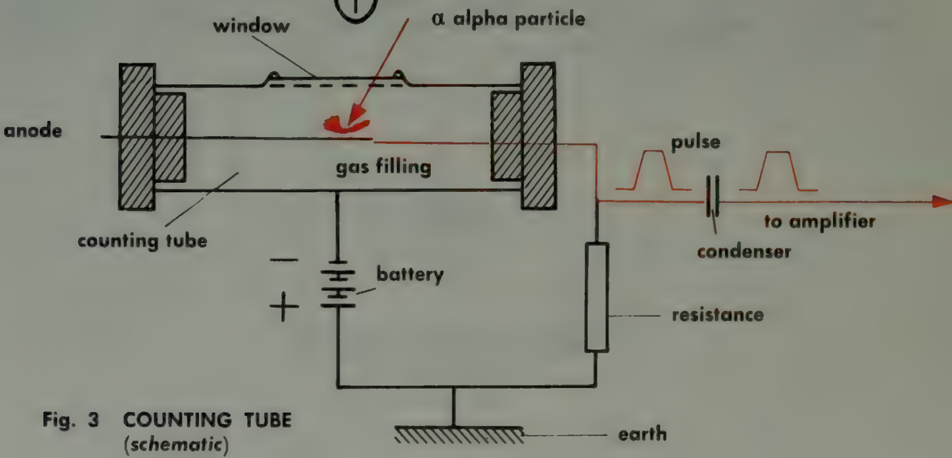


Fig. 3 COUNTING TUBE (schematic)

In a general sense the term "gyroscope" can be applied to any solid object rotating about a fixed point. For practical purposes a gyroscope consists of an axially symmetrical rotating body. It has a certain rotational momentum ("spin") which depends on the mass of the gyroscope, the square of the distance of the individual particles of mass to the axis of rotation, and on the speed of rotation (number of r.p.m.). To increase the rotational momentum, the gyroscope can advantageously be constructed as a disc with a thickened rim, so that most of its mass is concentrated as far from the axis as possible. The significant feature of a gyroscope is that the momentum and the rotational axis preserve their direction so long as no external forces act upon the gyroscope. Because of this tendency to keep the direction of its axis constant in space, the gyroscope can suitably be used for the stabilisation of movements. A convenient form of gyroscope (spinning top), in which the point of rotation coincides with the centre of gravity, is illustrated in Fig. 1a.

In the diagrams, forces are represented by black arrows, rotational momentum (N) and turning moments (M) by red arrows. The red arrows actually represent the so-called vector of the momentum or the turning moment; it can be conceived as the axis about which the momentum or moment rotates, the rotation being seen clockwise on looking towards the tip of the arrow. In the case of the rotational momentum (spin) the vector obviously coincides with the axis of rotation of the gyroscope itself. Now the behaviour of a gyroscope conforms to the following law: when a rotational moment is applied to a spinning gyroscope, the vector of the rotational momentum (and therefore the axis of the gyroscope) will tend to move in the direction of the vector of that applied moment. This can be explained with reference to Fig. 1b, where a gyroscope is shown tilted over to the left. Its own weight thus produces an overturning moment, which is in fact an applied moment in the sense envisaged above. This moment (M) has a vector represented by the thick red arrow, which must be conceived as pointing towards the viewer, i.e., perpendicular to the plane of tilting. It is in the direction of this arrow that the axis of the spinning gyroscope will swing. The axis thus swings round and round in the manner familiar to any one who has observed a child's spinning top.

The gyroscope effect can be demonstrated on a wheel mounted on a spindle. When it is spun round (Fig. 2a) and an attempt is made to tilt the wheel to the left—i.e., a turning moment to the left is applied—the momentum vector will tend to move in the direction of the thick red arrow (pointing forward). Similarly, if it is attempted to swivel the wheel to the left (Fig. 2b), the momentum vector will tend to swing upwards (so that the axis of the wheel tends to move in a vertical plane, whereas in Fig. 2a it tends to move in a horizontal plane). In ballistics the spin imparted to the projectile from a gun by the rifling (spiral grooves) of the barrel keeps it steady in flight. Some ships are equipped with large gyroscopes (Fig. 3) which act as stabilisers by damping the rolling movements.

Fig. 1a

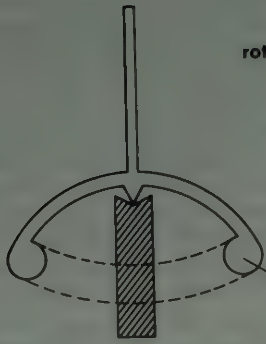
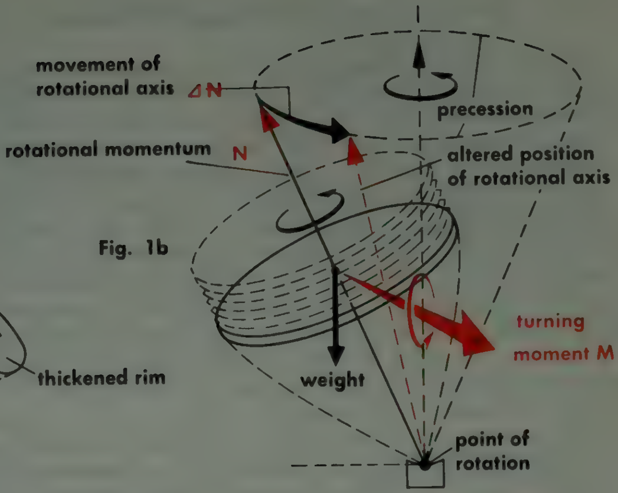


Fig. 1b



ig. 2a

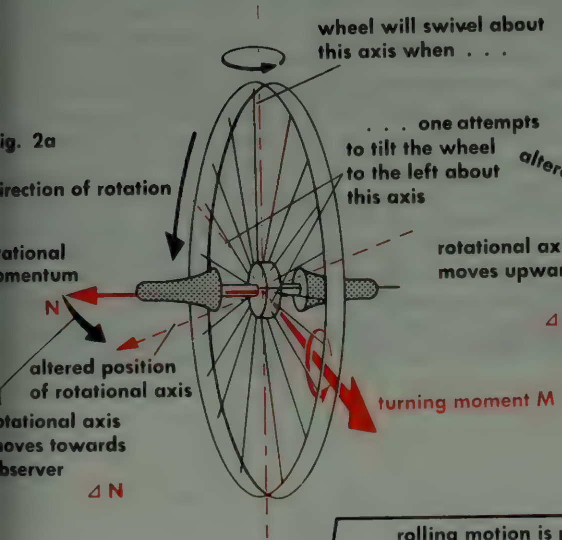
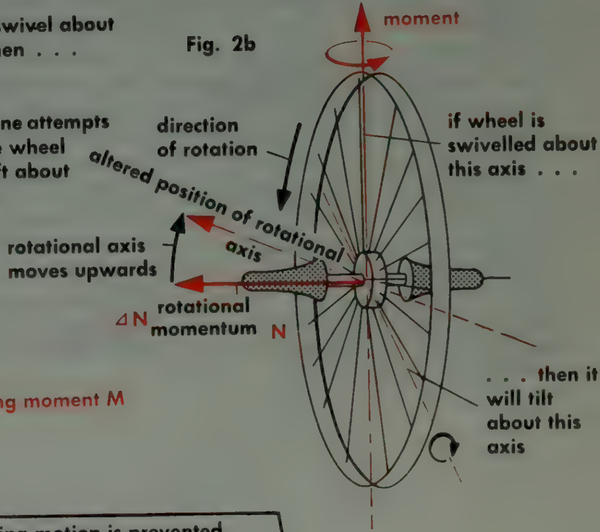


Fig. 2b



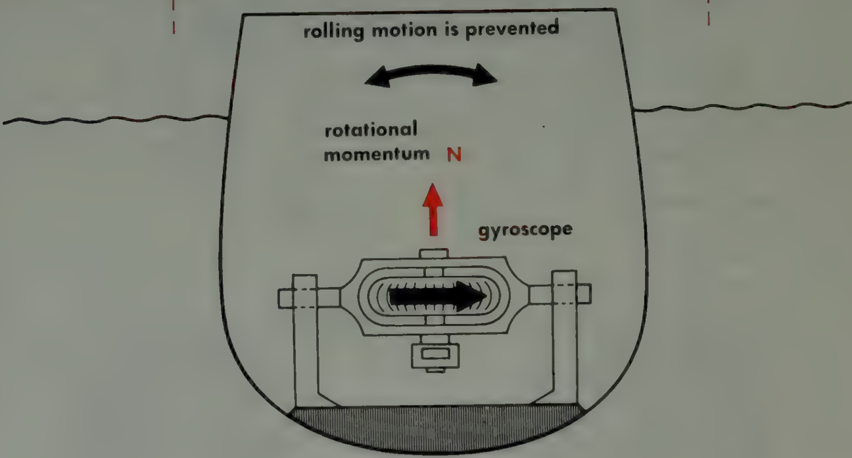


Fig. 3 GYRO-STABILISER

The freewheeling hub performs three functions: transmission of the driving power via the chain to the rear wheel in the course of normal forward motion of the pedals; freewheeling when the pedals are at rest; braking when the pedal motion is reversed.

The first two functions are performed by bringing a series of drive rollers into contact with a rotating cylinder and by releasing them respectively (Fig. 1). These rollers are disposed inside a guide ring. The drive wheel is fixed to a special ratchet which is so designed that when the pedals move forward (Fig. 1a) the sloped faces on the ratchet bear against the rollers so that the latter are jammed between the ratchet and the hub sleeve. In this way the power is transmitted through the rollers to the sleeve, which is fixed to the rear wheel of the bicycle and thus drives it round. When freewheeling, the ratchet remains stationary while the rear wheel continues to rotate (Fig. 1b). The rollers are then in the depressions formed in the surface of the ratchet and are no longer jammed against the hub sleeve. The latter is thus able to rotate freely in relation to the ratchet.

When the cyclist "back-pedals", the brake comes into action. The braking mechanism functions as follows. Two more or less conical components are slid into a slotted cylinder (the brake sleeve), causing this to expand, so that it is forced against the inside of the hub sleeve of the wheel, thus developing the braking action (Fig. 2). The inside of the roller guide ring is provided with two claws (Fig. 3) which hold the braking cone when back-pedalling takes place. The brake cone thus remains stationary, slides into the brake sleeve and, at the same time, pushes this sleeve on to the lever cone. It is this action that causes the expansion of the brake sleeve from both ends. The lever cone is provided with a lever (externally visible) secured to the frame of the bicycle and thus develops the reaction to the braking force applied. To do this, the lever cone must be prevented from rotating, and for this purpose it is provided with two flat surfaces (Fig. 5). Between each of these surfaces and the grooved inner surface of the brake sleeve is a roller, held in position by a spring. When the pedals are moved backwards, one end of each flat surface presses the roller into a groove, with the result that the cone is immovably gripped. The thrust exerted by the claws on the roller guide ring then forces the brake-sleeve over the lever cone. Fig. 4 shows a section through the hub.

Fig. 1

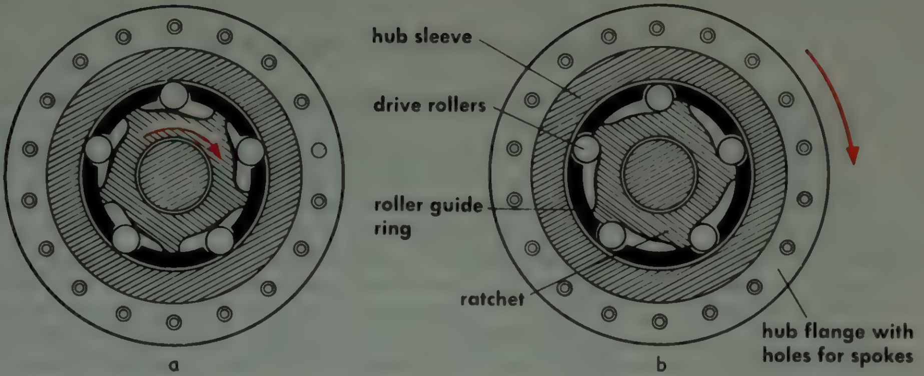


Fig. 2

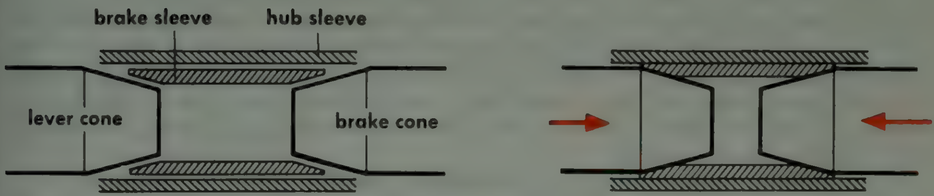


Fig. 3

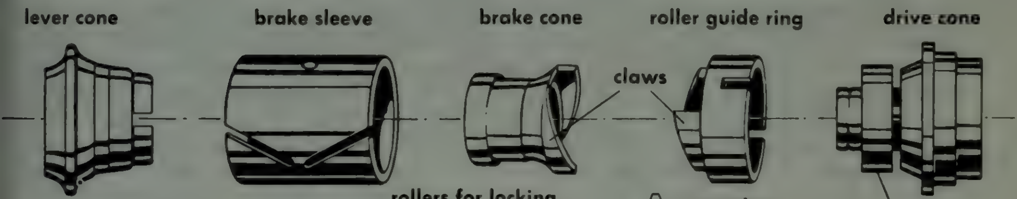


Fig. 4

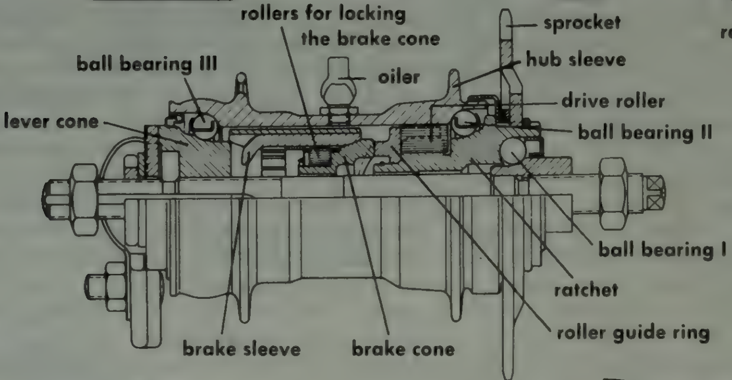
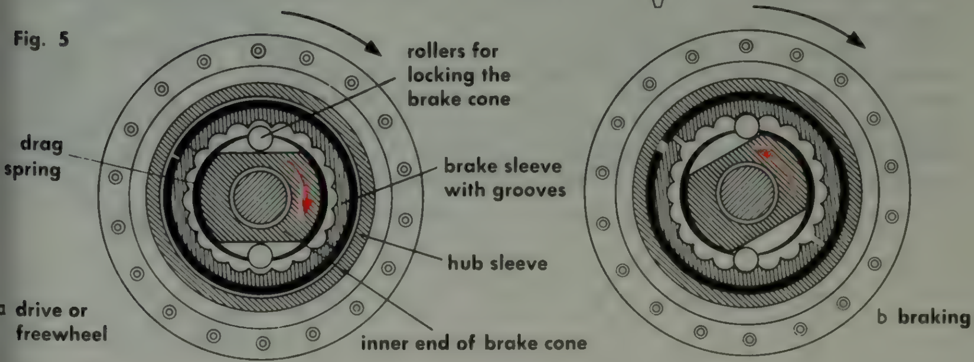


Fig. 5



There are two different systems of changing the gear ratios on a bicycle: the variable gear hub and the derailleur system.

The *variable gear hub* embodies a planetary (or epicyclic) gear set (Fig. 3) comprising a central cog-wheel, called the sun wheel, surrounded by three other cog-wheels called planet wheels. The latter engage with the sun wheel and also with internal gear teeth on a gear ring (annulus). The principle is illustrated in Figs. 1a and 1b, with reference to a system comprising only one planet wheel (the additional planet wheels in no way alter the principle). If the sun wheel is held stationary and the line connecting the centre of this wheel to the centre of the planet wheel is swung through a quarter circle, the planet wheel will rotate along the circumference of the sun wheel (with whose teeth it meshes) and will, in so doing, drive the annulus through a distance of more than a quarter circle (so that point A of the annulus thus arrives at A in Fig. 1b). The annulus thus rotates faster than the connecting line. Conversely, if the annulus is rotated, the connecting line will rotate at a lower number of revolutions than the annulus. The actual transmission ratio will, of course, depend on the relative dimensions and the numbers of teeth on all the gears involved. If the planet wheel is locked so that it is prevented from rotating about its own axis and the sun wheel is allowed to rotate about its axis, then the connecting line and the annulus will revolve at the same speed.

Fig. 2 shows a section through a three-speed variable gear hub. This hub has a planetary gear set for low gear (hill climbing gear) and another for high gear; both these gear sets engage with the same annulus. Gear-changing is effected by means of a gear control chain worked by the control wire from a lever mounted on the handlebars. The chain shifts the sun wheels of the two planetary gear sets along the hub spindle. When the sun wheel (which is additionally provided with internal teeth, Fig. 3) is slid over corresponding teeth on the stationary hub spindle and thus immovably locked in relation to the planet wheels, the transmission ratios envisaged above (i.e., the step-up ratio and the step-down ratio respectively) can operate. For "normal" gear, the internal teeth of the sun wheel are disengaged so that this wheel is now free to rotate on the hub spindle; at the same time the outer teeth of the sun wheel (which remain engaged with those of the planet wheels) are engaged with internal teeth of the cage which carries the three planet wheels (transmission ratio 1:1). The sprocket, driven by the chain from the pedals, is fixed to the planet wheel cage (the outer drive element) of one of the planetary gear sets. When the gear set shown on the right in Fig. 2 is engaged, the power is transmitted from the driven cage through the faster-rotating annulus to the left-hand gear set—which is now locked—and thus to the planet wheel cage (the inner drive element) of this left-hand gear set. The inner drive element is provided with a worm thread on which the drive cone runs. When this cone is moved over to the right, it carries along with it the surrounding cone of the drive sleeve. The latter is fixed to the hub and thus to the rear wheel of the bicycle. In its left-hand position the drive cone performs a braking function (see page 182). This is the situation for high gear. In low gear (for hill climbing), the gear set on the left is operative, while the right-hand set is locked instead. In "normal" gear both planetary gear sets are locked, and the drive is thus directly transmitted from the sprocket to the rear wheel of the bicycle.

A *derailleur gear* system comprises a freewheel with two or more sprockets, a mechanism that alters the line of the chain and causes it to jump from one sprocket to another, and a spring-operated jockey pulley or tension pinion to take up or let out the slack in the chain. In this way various gear ratios can be selected. Although the principle is the same in all derailleur gears, the mechanisms produced by the various manufacturers differ in detail. A typical derailleur is illustrated in Fig. 4.

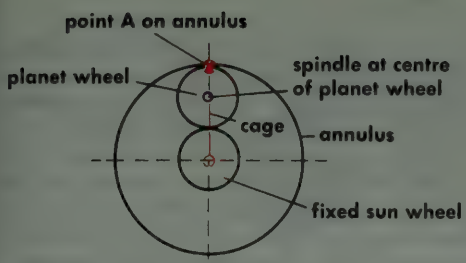


Fig. 1a PLANETARY GEAR SET WITH FIXED SUN WHEEL

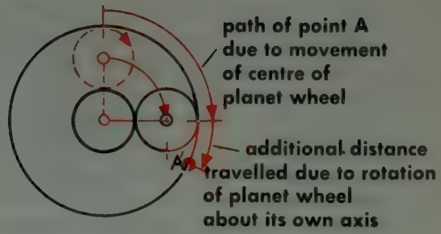


Fig. 1b POSITION AFTER A QUARTER TURN

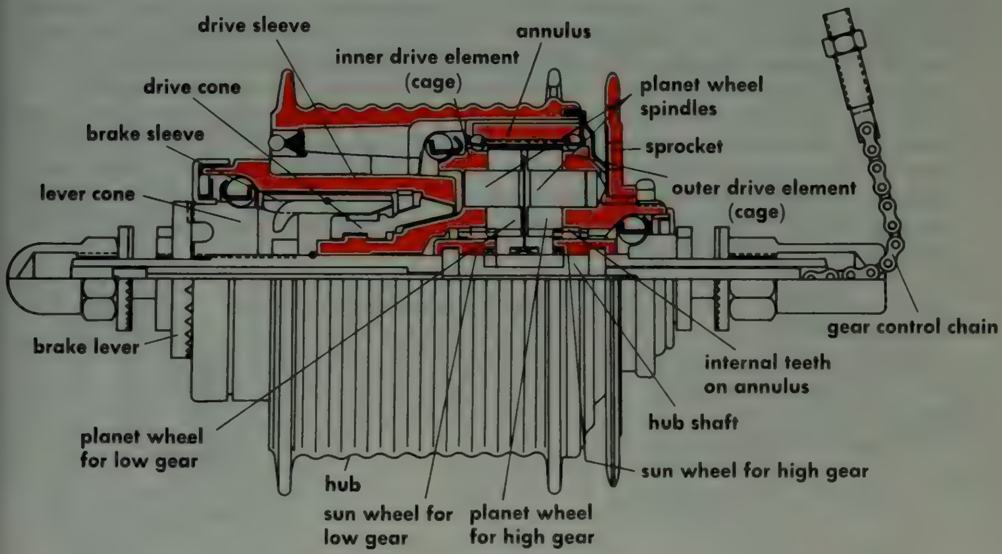


Fig. 2 SECTION THROUGH A THREE-SPEED VARIABLE-GEAR HUB
Setting: "normal" gear

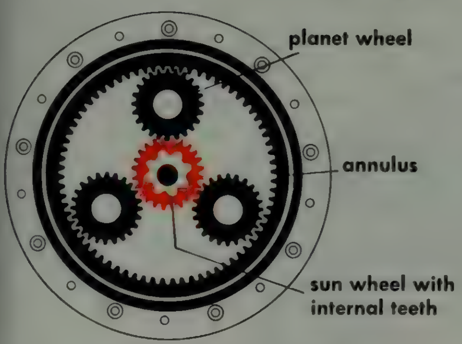


Fig. 3 PLANETARY GEAR SYSTEM OF THREE-SPEED VARIABLE-GEAR HUB

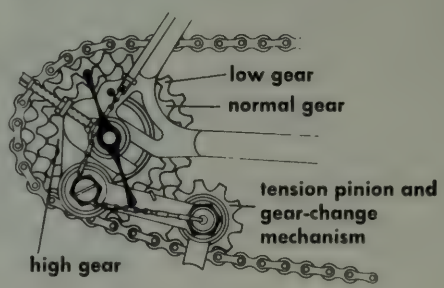


Fig. 4 DERAILLEUR GEAR

Two-wheeled vehicles powered by internal combustion engines comprise motor cycles, motor scooters and mopeds.[1] These are all similar in principle.

The motor cycle comprises four main sections: the frame, the engine with gearbox and drive components (chain or drive shaft), the road wheels, and the petrol[2] tank (Fig. 1). The frame is of tubular pressed steel construction. The most important components of the frame are the springs for absorbing the jolts caused by the irregularities of the road surface or other obstacles encountered. There are various kinds of springing for motor cycles: (1) parallelogram suspension (Fig. 2), which is a somewhat clumsy system because the whole mass of the front wheel and fork participates in the oscillation; (2) telescopic front forks in combination with short swinging arm (Fig. 3), in which arrangement the oscillating mass is reduced; (3) telescopic front forks in combination with long swinging arm (Fig. 4): more resilient than the foregoing, but involving great oscillating mass (the same principle is applied to the rear wheel suspension in Fig. 3). The telescopic suspension system used in types (2) and (3) has the advantage that it can undergo considerable compression in conjunction with only small oscillating masses. By these or similar means it is possible to achieve very efficient suspension conditions, so that the saddle springing can be dispensed with on modern motor cycles. A foam-rubber seating cushion is usually sufficient, especially on racing machines. Sectional views of typical front and rear suspension systems are shown in Figs. 4 and 5. Spring elements in combination with oil-filled cylinders provide the necessary shock-absorbing function (see page 232).

Two-stroke and four-stroke engines are used as power units for motor cycles. The power is transmitted to the rear wheel through the gearbox (Fig. 6) and thence through sprockets and chains or through a drive shaft. Transmission through a shaft is shown in Fig. 5, the shaft being neatly accommodated in the swinging arm of the rear suspension. Motor cycles and mopeds have wire-spoke wheels, whereas scooters generally have solid wheels like those of a car.

1. "Moped" is an abbreviation of "motor-assisted pedal cycle".
2. Gasoline in U.S.A.

Fig. 1 MOTOR CYCLE

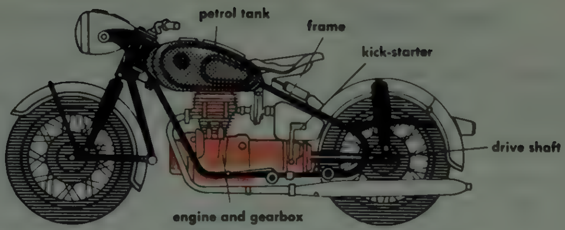


Fig. 2 PARALLELOGRAM SUSPENSION

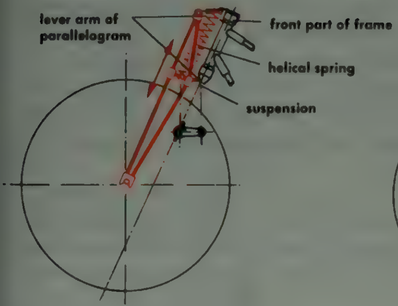


Fig. 3 FRONT AND REAR WHEEL SUSPENSION WITH SWINGING ARMS AND TELESCOPIC SHOCK ABSORBERS

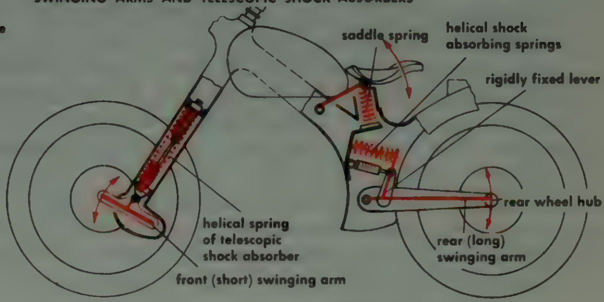


Fig. 4 FRONT WHEEL SUSPENSION WITH TELESCOPIC SHOCK ABSORBER AND LONG SWINGING ARM

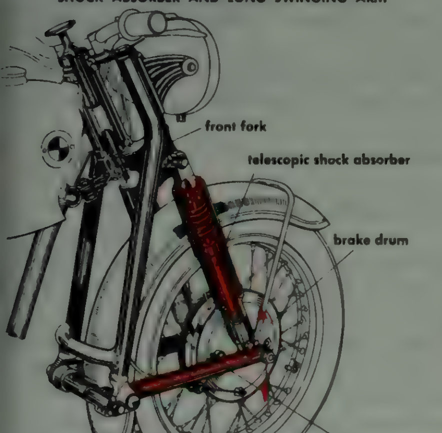


Fig. 5 REAR WHEEL WITH DRIVE, BRAKE AND SUSPENSION ELEMENTS

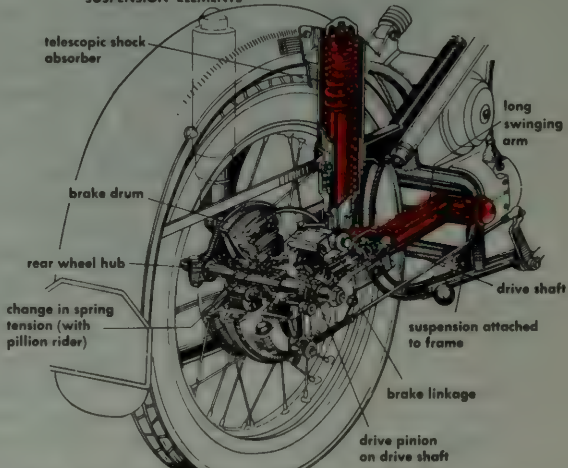
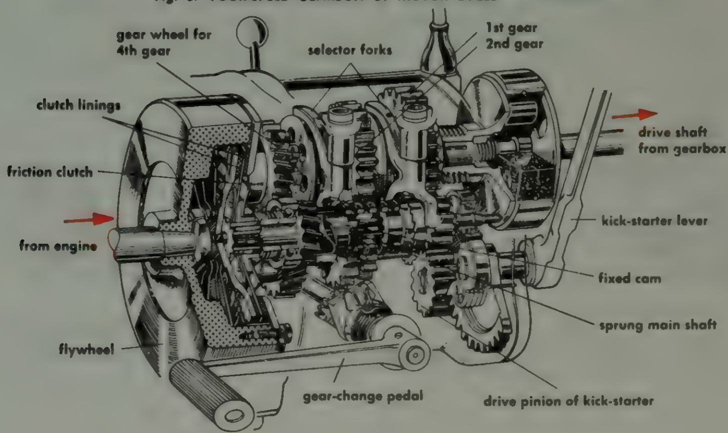


Fig. 6. FOUR-SPEED GEARBOX OF MOTOR CYCLE



INTERNAL COMBUSTION ENGINE: FOUR-STROKE PETROL (GASOLINE) ENGINE

The petrol engine, like the diesel engine (page 194), is an internal combustion engine. The thermal energy which is released when the fuel is burned is converted into mechanical energy. The petrol engine differs from the diesel in that the liquid fuel (i.e., the petrol) is mixed with air—usually in a device called a carburettor (see page 202)—to form a combustible mixture, which is compressed in the cylinder and finally ignited by an electric spark produced between the electrodes of a sparking plug (see page 208). The gases which are formed in the cylinder by the combustion of the petrol-and-air mixture expand and thrust the piston downwards. Acting through the connecting rod, the piston imparts a rotary motion to the crankshaft. The spent burned gases must then be removed from the cylinder and be replaced by fresh petrol-and-air mixture, so that a fresh cycle can begin. The energy needed for effecting this change in the contents of the cylinder is provided by the flywheel, which stores up some of the mechanical energy released by the combustion that takes place in the cylinder. The additional energy developed by the engine can be taken off at the end of the crankshaft.

With internal combustion engines—diesel as well as petrol engines—a distinction must be made between four-stroke and two-stroke operation. To perform a full cycle of operations (changing the contents of the cylinder and effecting the combustion) the four-stroke engine requires four, and the two-stroke engine requires two strokes of the piston.

Four-stroke engine:

1st stroke: induction stroke: while the inlet valve is open, the descending piston draws fresh petrol-and-air mixture into the cylinder.

2nd stroke: compression stroke: While the valves are closed, the rising piston compresses the mixture to a pressure of about 7–8 atm.; the mixture is then ignited by the sparking plug.

3rd stroke: power stroke: While the valves are closed, the pressure of the gases of combustion forces the piston downwards.

4th stroke: exhaust stroke: the exhaust valve is open and the rising piston dis charges the spent gases from the cylinder.

Since power is developed during one stroke only, the single-cylinder four-stroke engine has a low degree of uniformity, i.e., the rotation of the crankshaft is subject to considerable accelerations and decelerations during a cycle. More uniform—that is to say, smoother—running is obtained with multi-cylinder engines because the "cranks" of the crankshaft are staggered in relation to one another (Fig. 3), so that the various cylinders do not develop their power strokes simultaneously, but successively (and sometimes in an overlapping sequence). Depending on the cylinder arrangement, various types of engine are to be distinguished: in-line engine (Fig. 3), horizontally opposed engine (Fig. 4), vee engine (Fig. 5), and radial engine (Fig. 6).

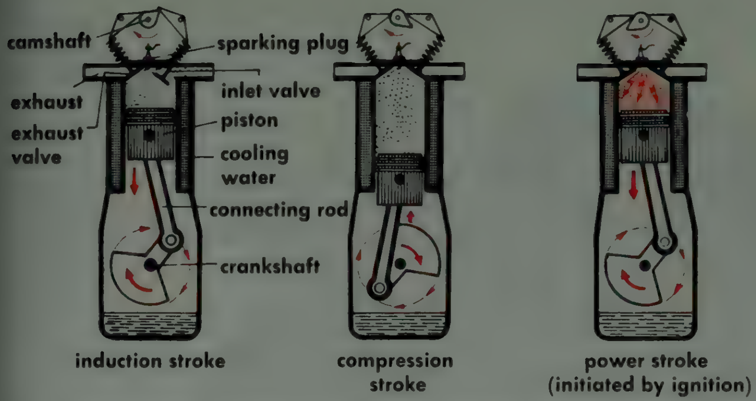


Fig. 1 OPERATING PRINCIPLE OF A FOUR-STROKE ENGINE

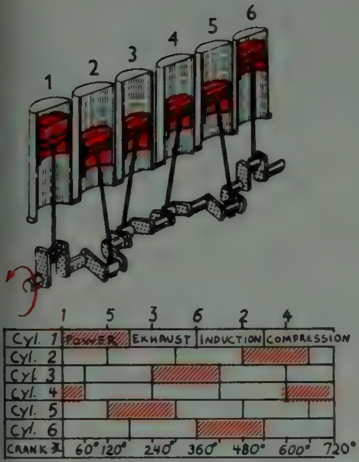


Fig. 2 WORKING SEQUENCE OF A SIX-CYLINDER IN-LINE ENGINE

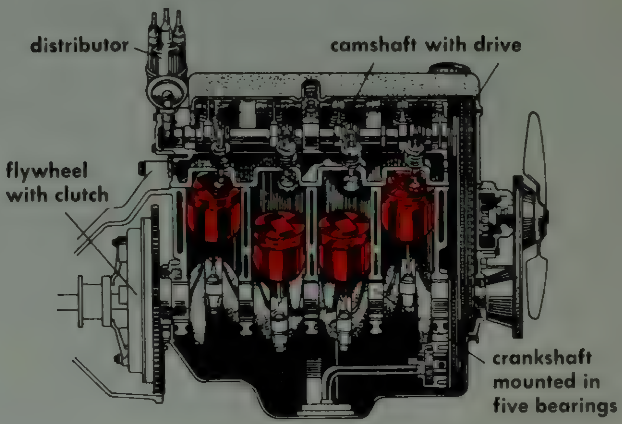


Fig. 3 SECTION THROUGH A FOUR-CYLINDER IN-LINE ENGINE

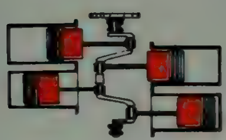


Fig. 4 HORIZONTALLY OPPOSED ENGINE

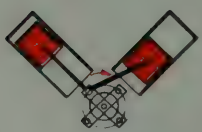


Fig. 5 VEE ENGINE

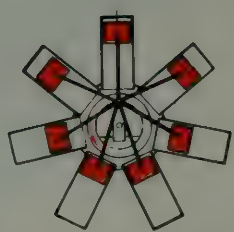
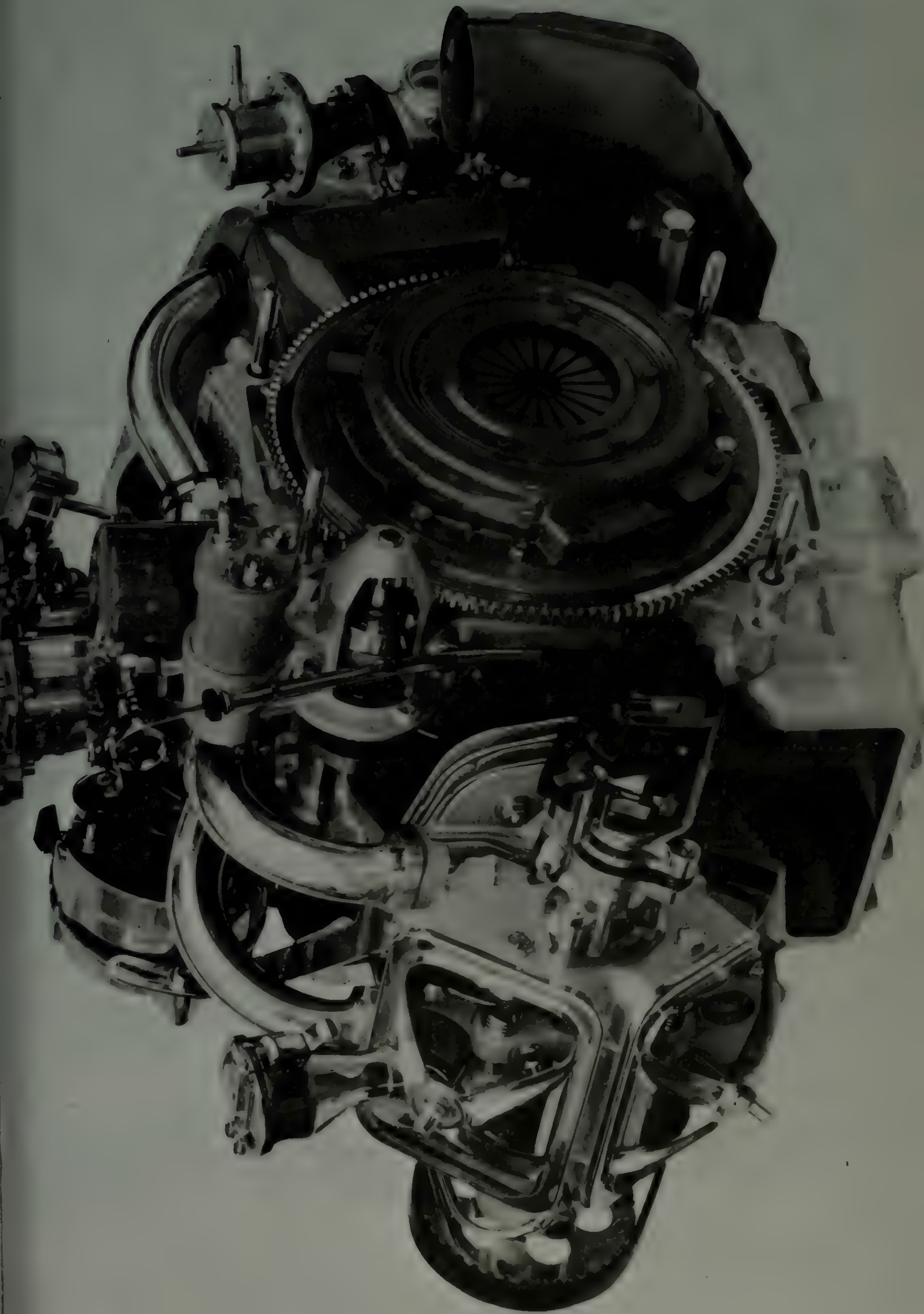


Fig. 6 RADIAL ENGINE

Citroën G.S. four-stroke engine
Photo Citroën



INTERNAL COMBUSTION ENGINE: TWO-STROKE PETROL (GASOLINE) ENGINE

In this type of engine the piston periodically covers and uncovers openings—known as ports—in the cylinder wall (the two-stroke engine is seldom equipped with valves).

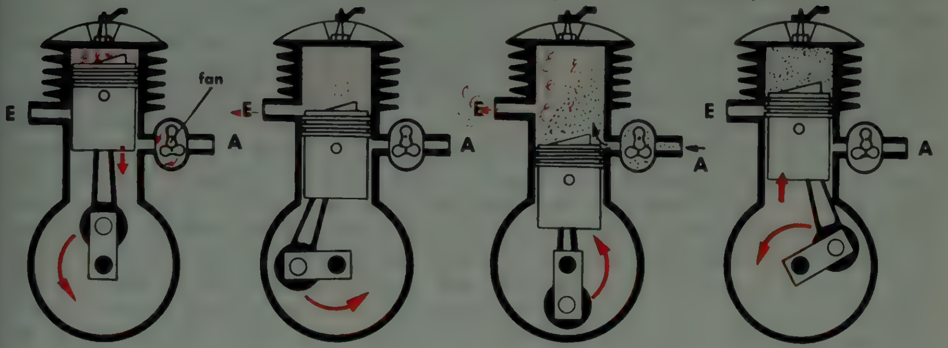
The operation of the two-stroke engine will be explained with reference to Fig. 7, which represents such an engine of the type provided with a scavenging fan. At the start of the first stroke, the piston is in its highest position. When the compressed petrol-and-air over the piston is ignited, the latter is thrust downwards and, in so doing, releases the exhaust port. The burned gases in the cylinder, which are still under high pressure, can thus escape through this port. When the piston descends further, its upper edge releases the inlet port, which admits fresh petrol-and-air mixture (delivered by the fan) into the cylinder, so that the remaining burned gases are flushed out. When the piston rises again (2nd stroke), all the ports are closed for a time, and during this period the petrol-and-air mixture is compressed, so that a fresh cycle can commence.

The crankcase-scavenged two-stroke engine (Fig. 8) has no scavenging fan. Instead, the crankcase is hermetically sealed, so that it can function as a pump in conjunction with the piston. When the piston ascends, a partial vacuum is produced in the crankcase, until the lower edge of the piston releases the inlet port and thus opens the way to the fresh petrol-and-air mixture into the crankcase. When the piston descends, the mixture in the crankcase is compressed a little so that, as soon as the top of the piston releases the transfer port and overflow duct (connecting the crankcase to the cylinder), it can enter the cylinder. Meanwhile, what happens above the piston is the same as in the fan-scavenged engine.

In the latter type of two-stroke engine the fan adds to the cost. However, as the overflow duct between the cylinder and crankcase is eliminated, the crankshaft can be provided with forced-oil lubrication without involving a risk that the oil in the crankcase can find its way into the cylinder. In the cheaper crankcase-scavenged engine the lubricating oil is mixed with the petrol ("petroil" lubrication) or is, alternatively, supplied to the points of lubrication dropwise by small lubricating oil pumps. The oil which enters the crankcase is liable to be carried through the overflow duct and transfer port into the cylinder, whence it passes through the exhaust port and into the exhaust system, where it may manifest itself as blue smoke in the exhaust.

Fig. 7 FAN-SCAVENGED TWO-STROKE ENGINE

A = admission port (*inlet port*) E = exhaust port



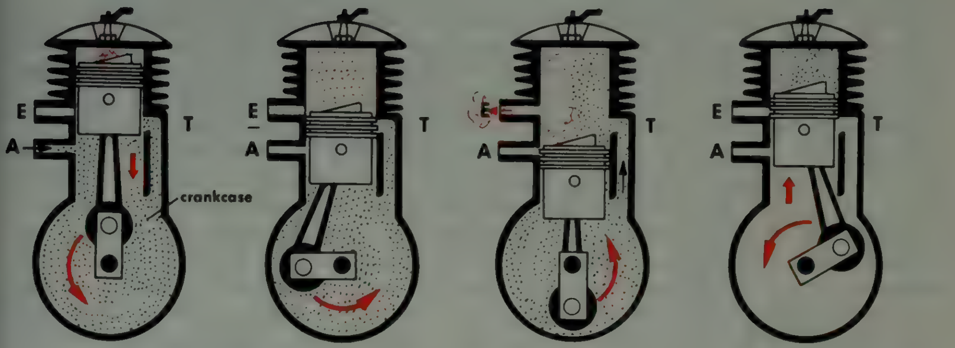
after ignition, power is developed as the piston descends

when the exhaust port E is open, the burned gas escapes

when the admission port A is open, the fan forces fresh petrol-and-air mixture into the cylinder

when the piston rises and E and A are closed, compression takes place

Fig. 8 CRANKCASE-SCAVENGED TWO-STROKE ENGINE



above the piston

after ignition, power is developed as the piston descends

when the exhaust port E is open, the burned gas escapes

when the transfer port T is open, petrol-and-air mixture flows from the crankcase (where it has been compressed) into the cylinder

when the piston rises and E and T are closed, compression takes place

below the piston

when the admission port A is open, fresh petrol-and-air mixture flows into the crankcase

when E and T (transfer port) are closed, petrol-and-air mixture in the crankcase is compressed

after A and T are closed, further rising of the piston produces suction in the crankcase

The combustion processes in the petrol[1] engine and the diesel engine differ in the following significant features: in the petrol engine (see page 188) the petrol-and-air mixture is drawn into the cylinder, compressed (compression ratio ranging from 4:1 to 10:1), and ignited by a spark. In the diesel engine, on the other hand, air alone is drawn into the cylinder and is compressed to a much higher ratio (14:1 to 25:1) than in the petrol engine. As a result of this high compression the air is heated to a temperature of 700°–900° C. Only then is a certain quantity of diesel fuel injected into the cylinder. Beeause of the prevailing high temperature, the fuel ignites spontaneously. However, combustion does not take place immediately when the fuel particles enter the combustion chamber, but after an interval of about $\frac{1}{1000}$ sec. This is because the fuel droplets first have to mix intimately with the air in the combustion chamber and must then be heated up and vaporised before they can burn. The time that elapses between injection and ignition is called the ignition lag. On injection of the fuel, it is broken up by the nozzle into smaller and larger droplets, according to a certain pattern. The smaller droplets occur more particularly in the edge zone of the injected fuel spray (Fig. 1) and are the first to ignite. Next, the larger droplets in the interior of the spray are ignited. Fuel injection continues after the first flame has formed (main combustion). If some of the diesel fuel is incompletely burned in this combustion process, or if it accumulates and then burns suddenly and violently at the next main combustion, the engine is said to be "knocking". In the petrol engine, on the other hand, the petrol-and-air mixture first ignites in the vicinity of the sparking[2] plug (Fig. 2). The heat given off by the burning fuel particles causes the adjacent particles to ignite, so that a flame front, starting from the sparking plug, spreads through the combustion chamber. As a result of thermal radiation and rise in pressure, but also in consequence of the presence of "hot pockets" in the combustion chamber, a fresh ignition and flame front formation may occur in the still unburned mixture not yet reached by the initial flame front. Thus, in certain circumstances, the entire unburned mixture may undergo sudden and violent combustion, causing "knocking" of the petrol engine.

Diesel engines are designed to operate on the four-stroke or the two-stroke principle, just like petrol engines. In respect of their design features diesel engines differ more particularly in the manner in which the fuel and air are brought together and ignited. In a diesel engine with direct injection (or solid injection) the fuel is injected directly into the cylinder (or sometimes into a spherical combustion chamber formed in the piston) (Fig. 3). The injection nozzle must be so designed that its spray pervades all the air in the combustion chamber, so that complete combustion of the fuel can take place. In a diesel engine provided with a swirl chamber (Fig. 4) the air is forced into this chamber by the piston and thereby acquires a rapid swirling motion, so that mixing with the fuel is promoted. What mainly occurs in the precombustion chamber of a diesel engine as shown in Figs. 5 and 6 is mixing of the fuel and air and preliminary combustion of the mixture. As a result of the rise in pressure during this precombustion, the incompletely burned fuel-and-gas mixture flows at high velocity from the precombustion chamber into the cylinder, where it can undergo complete combustion with the air there.

1. Gasoline in U.S.A.
2. Spark plug in U.S.A.

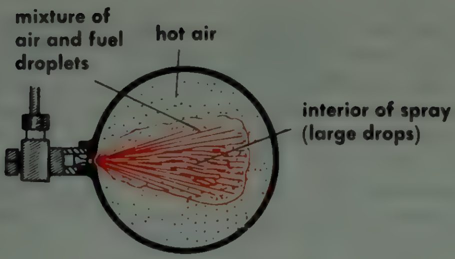


Fig. 1 COMBUSTION PROCESS IN A DIESEL ENGINE

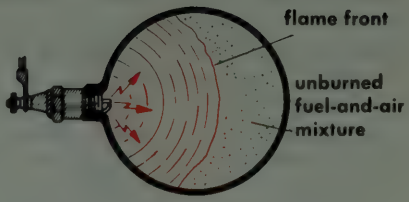
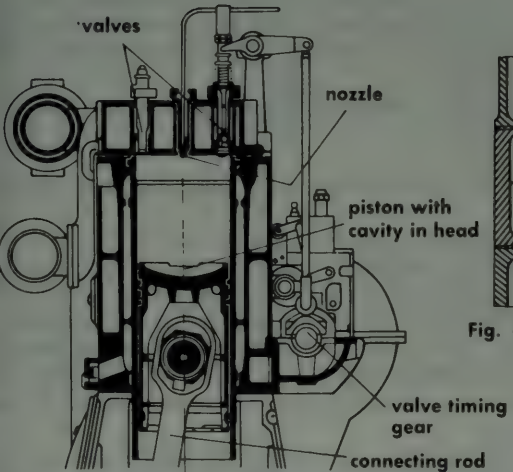


Fig. 2 COMBUSTION PROCESS IN A PETROL ENGINE



3 DIESEL ENGINE WITH DIRECT INJECTION

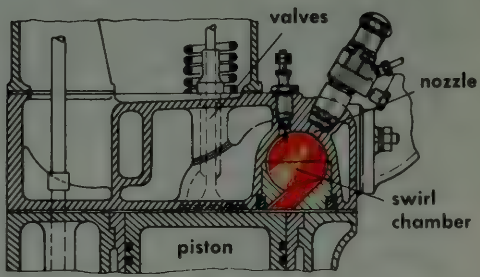


Fig. 4 DIESEL ENGINE WITH SWIRL CHAMBER

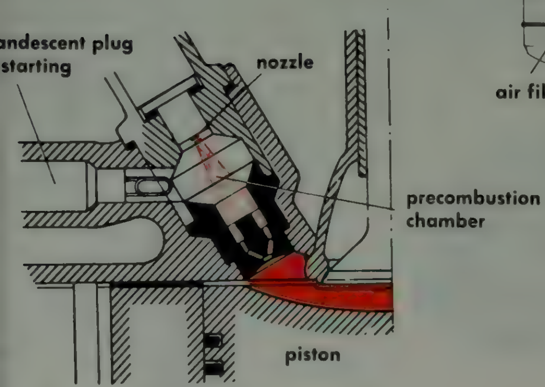


Fig. 5 PRECOMBUSTION CHAMBER OF A DIESEL ENGINE

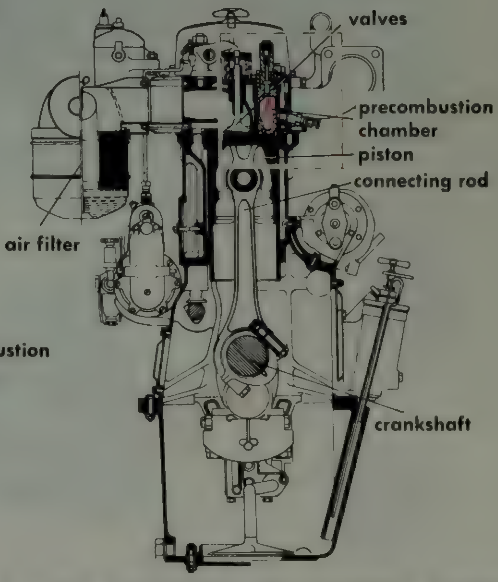


Fig. 6 DIESEL ENGINE WITH PRECOMBUSTION CHAMBER

The rotary piston engine is an internal combustion engine which operates on the same general principle as the conventional petrol[1] engine (see page 188). In the latter, however, a rotary motion is produced by an oscillating piston and connecting rod (Fig. 1), whereas the rotary piston engine produces this motion by means of a rotating "piston". This means that there are no oscillating masses which have to be alternately accelerated and retarded, as occurs when an ordinary piston moves to and fro. Consequently, the forces of inertia associated with the oscillatory motion are obviated in the rotary piston engine. As a result, higher speeds of rotation are possible. The edges of the rotating piston open and close ports in the cylinder wall, so that the piston itself controls the "breathing" of the engine, without the aid of valves. The triangular piston with convex sides rotates in a housing whose internal cross-section presents an oval shape slightly constricted in the middle (epitrochoid). When the piston rotates, the seals mounted at its three corners continuously sweep along the wall of the housing. The three enclosed spaces formed between the piston and the wall successively increase and decrease in size with each revolution. These variations in the spaces are utilised for drawing in the fuel-and-air mixture, for compressing this mixture, for combustion, and for discharging the burned gases, so that the full four-stroke working cycle is performed (Fig. 2). The four strokes for the chamber between the corners A and C are as follows:

1st stroke: the rotary piston uncovers the inlet port: the mixture of fuel and air flows in;

2nd stroke: the mixture is compressed;

3rd stroke: the compressed mixture is ignited by the sparking plug, burns and drives the piston (power stroke);

4th stroke: the exhaust gas is discharged through the outlet port.

In the other two combustion chambers the same processes occur, but with a displacement of 120° in each case. In the course of one piston rotation there are therefore three ignitions. As appears from Fig. 2, the rotary piston engine has only two moving parts: the piston and the driving shaft. The piston is provided with concentric internal gearing which engages with a gear wheel which is concentric with the engine shaft but is firmly connected to the housing. The ratio of the number of teeth of the internal gearing and gear wheel is 3:2.

One of the major problems in the construction of the rotary piston engine is the sealing of the three chambers in relation to one another. Intercommunication between these chambers would be detrimental to the proper functioning of the engine. This problem has been solved by means of a system of sealing strips (Fig. 4).

1. Gasoline in U.S.A.

Fig. 1 COMPARISON OF MODE OF OPERATION OF TWO-STROKE PISTON ENGINE AND ROTARY PISTON ENGINE

Fig. 2 SEQUENCE OF OPERATIONS IN THE ROTARY PISTON ENGINE
(shown for one combustion chamber)

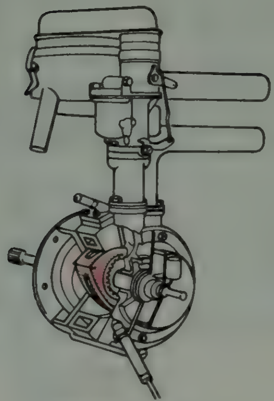


Fig. 3 SECTION THROUGH A ROTARY PISTON ENGINE

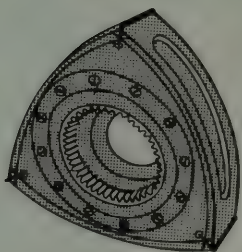


Fig. 4 ROTARY PISTON

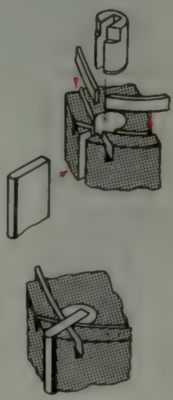
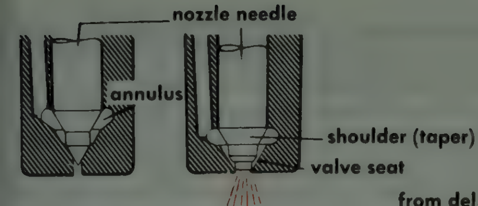


Fig. 5 PISTON EDGE SEAL

In a diesel engine (see page 194) the fuel-and-air mixture is formed within the cylinder itself, after compression by the piston. The fuel is injected into the highly compressed hot air and burns. The fuel injection pump has to provide the requisite high pressure (80–300 atm. = 1175–4400 lb./in.^{2}), which must be higher than the pressure developed in the cylinder at the instant of injection, and ensure that a certain quantity of fuel is injected into the combustion chamber at the correct time and with the desired velocity. The injection nozzle distributes and atomises the fuel (Fig. 1). It is mounted in the nozzle holder, and the nozzle needle is pressed against the valve seat by a spring. The needle has a larger cross-section at its top part than at the bottom. The pump pressure acts upon the transition surface (taper) between the two cross-sections (Fig. 1b), so that, when the spring pressure is overcome, the needle is raised some distance and the fuel is injected into the combustion chamber through the aperture thus formed. The nozzle holder is provided with a screw by means of which the spring pressure and therefore the injection pressure (i.e., the pressure needed to overcome the counteracting force exerted by the spring) can be varied. The nozzle holder is connected to the injection pump by a delivery pipe. The length of this pipe, and indeed its diameter and wall thickness, affect the pressure distribution in the pipe and thus also affect the injection and the combustion process.

Like every reciprocating pump (p. 30, vol. I) the injection pump comprises a piston, a suction valve and a delivery valve. In addition, it is provided with a device for regulating the quantity of fuel delivered. In general, this regulation is effected by keeping the suction valve open for a certain length of time while the piston is performing its stroke. As a result, a proportion of the delivered quantity is returned to the suction pipe. In Bosch pumps (Fig. 3) the delivery rate is varied by rotating the piston. The compression chamber over the piston is connected to the space below the chamber of the piston by a longitudinal groove or hole. So long as the surface shown in red in Fig. 4 covers the opening on the right, the pump delivers fuel; when the opening is uncovered, however, delivery ceases because then the rest of the fuel, which is displaced above the piston, can flow away through the longitudinal groove into the space below, whence it is discharged through the overflow pipe. On rotation of the piston, the "red" area will cover the opening for a longer or shorter time, so that the delivery rate is varied. This adjustment by rotating the piston is performed by hand (through the agency of a toothed rack) or by a centrifugal governor (Fig. 5).

In order to achieve a rapid drop in pressure in the delivery pipe from the pump to the nozzle, and thus to ensure immediate closure of the latter, the pressure valves of Bosch injection pumps are fitted with a special device (Fig. 6). On completion of delivery, the relief piston first plunges into the valve hole and closes the delivery pipe. Then the taper of the valve body is lowered on to a seat. As a result, the volume available to the fuel in the delivery pipe undergoes an increase equal to the volume of the relief piston, and the delivery pipe is quickly relieved of pressure. Dribbling and the resultant carbonisation (coking) of the nozzle are thereby obviated.



1a NEEDLE
ZZLE CLOSED

Fig. 1b PINTLE
NOZZLE OPEN

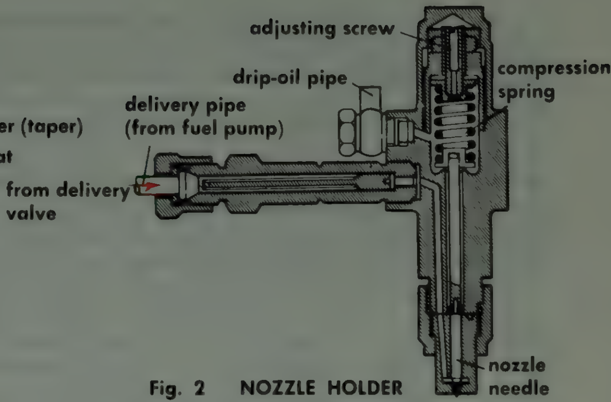
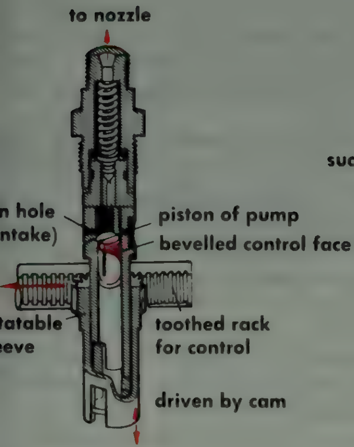


Fig. 2 NOZZLE HOLDER



FUEL INJECTION PUMP ELEMENT

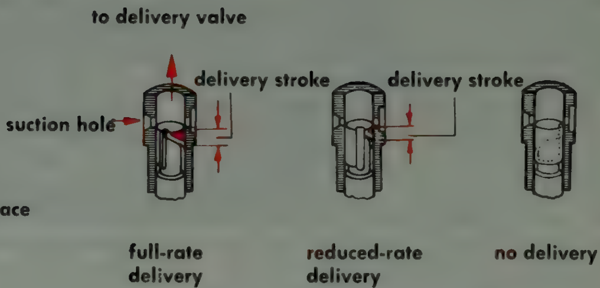
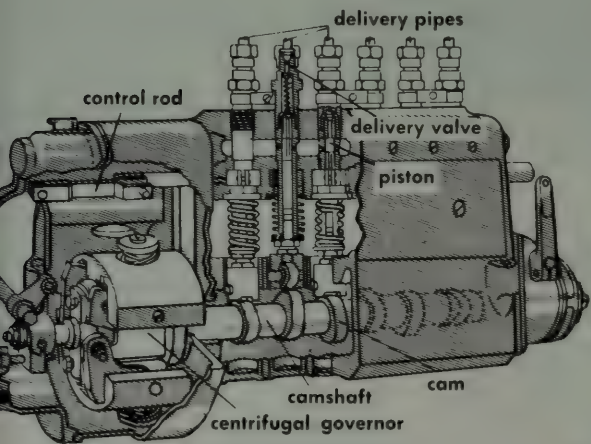


Fig. 4 ROTATION OF PISTON
VARIES DELIVERY RATE



g. 5 INJECTION PUMP (cut-away view)

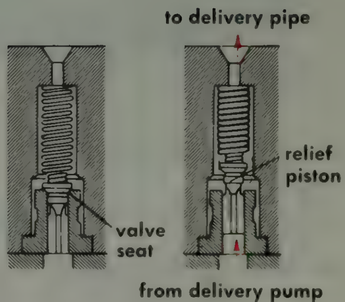


Fig. 6 PRESSURE VALVE WITH RELIEF PISTON

Internal combustion engines are generally started by rotating the flywheel until the engine fires and continues to run on its own power. The flywheel is usually rotated by a starter motor fed with current from the battery. Petrol[1] engines have to be started at a speed of 50–60 r.p.m.; diesel engines require about 100 r.p.m. The shaft of the starter is provided with a pinion (a small gear wheel) which, on commencement of the starting operation, is shifted forward until it engages with the toothed rim on the flywheel. The driving motor of the starter then rotates the flywheel. When the engine has started up, the starter pinion is disengaged from the flywheel and retracted. The various types of starter which are in present-day use differ mainly in the manner in which engagement and disengagement is effected.

Pinion shift starter (Figs. 1 and 2): When the starter button (inside the car) is actuated, the connection between the battery and the solenoid switch is established. A powerful magnetic field is set up in the magnetic coil of this switch, so that the armature of the switch is pulled forward such a distance that the tilting bridge closes the contact I, with the result that the auxiliary coil and the holding coil are energised. This causes the armature to be drawn, slowly rotating, into the magnetic field of the auxiliary coil and to engage with the toothed rim of the flywheel.

As a result of this movement the release lever is freed and closes the contact II on the tilting bridge (Fig. 1b). The main coil is now energised, the armature rotates, and the engine is turned. When the starter button is released, the current ceases to flow through the armature, and the latter is pulled back, and thus disengaged from the flywheel, by the return spring.

Inertia gear drive starter (Bendix type starter) (Figs. 3 and 4): In this type of starter the pinion can be shifted along the armature shaft on a quick screw thread. When the starter button is actuated, the armature of the solenoid switch is attracted and the pinion is pushed forward, while it rotates, by the engaging lever. In the second stage the solenoid switch closes, the current ceases to flow through the draw-in coil, the armature of the starter begins to rotate and screws the pinion completely forward until it engages with the gear rim on the flywheel. The solenoid switch is held in position by the holding coil. The starter motor can now turn the engine. When the latter fires, the engagement lever is pulled back to the home position by the return spring; when the engine begins to rotate faster than the starter, the pinion is disconnected from the armature shaft by a roller-operated freewheel device.

With all types of starter it is possible to perform the first stage of operation by means of a pedal, as exemplified by a gear-shift starter (Fig. 5). The design is simpler than in the case of the Bendix type starter because here the pinion is not rotated while it engages with the gear rim.

1. Gasoline in U.S.A.

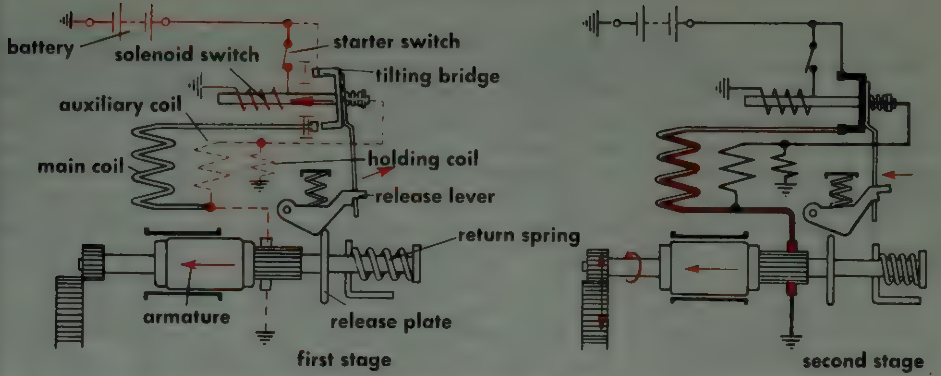


Fig. 1 OPERATING STAGES OF PINION-SHIFT STARTER

Fig. 2 PINION-SHIFT STARTER

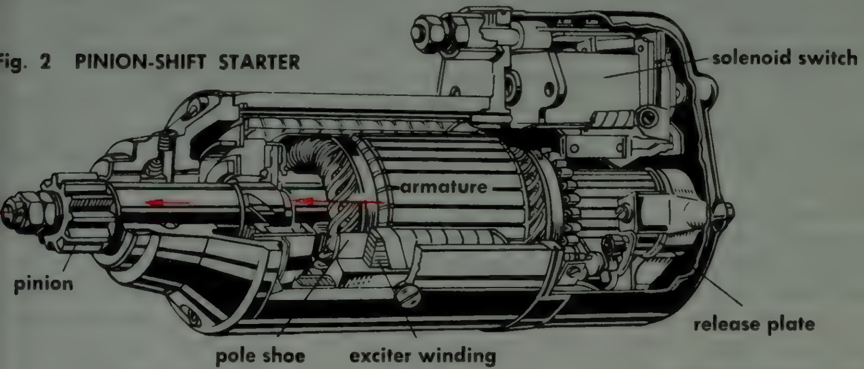
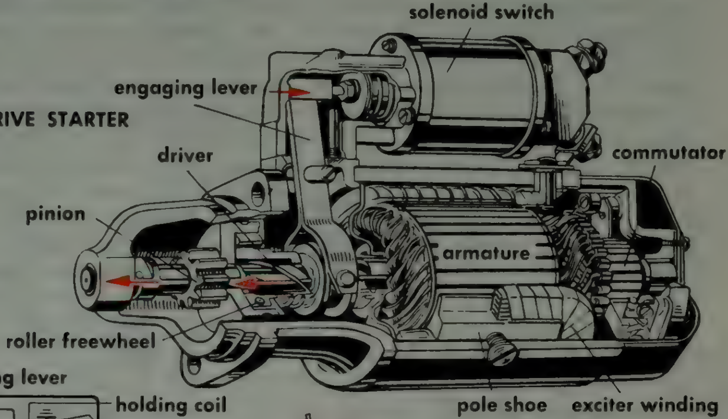


Fig. 3 INERTIA GEAR DRIVE STARTER



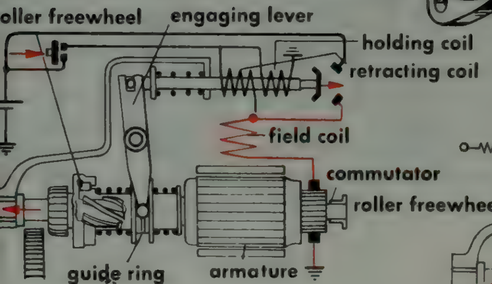


Fig. 4 INERTIA GEAR DRIVE STARTER

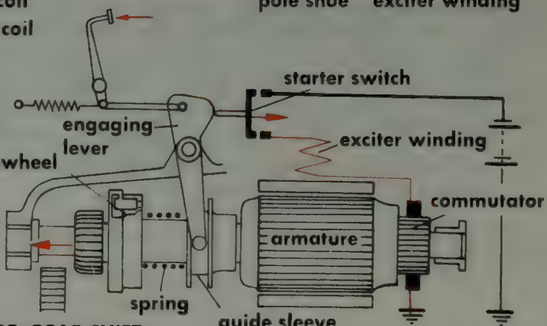


Fig. 5 SINGLE-STAGE GEAR-SHIFT STARTER

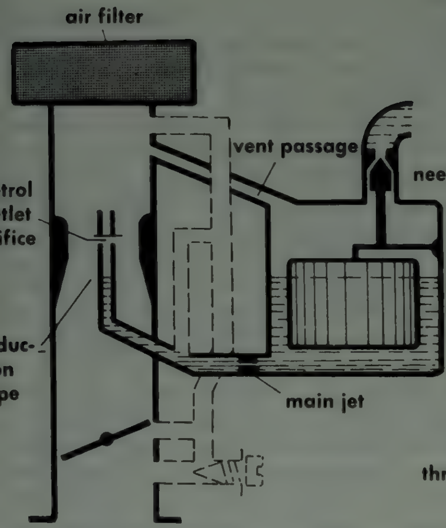
CARBURETTOR[1]

The function of the carburettor is to produce the fuel-and-air mixture needed for the operation of a petrol[2] engine. In the carburettor the fuel is distributed in the form of tiny droplets in the stream of air. As a result of heat absorption on the way to the cylinder these droplets are vaporised, so that the mixture thus becomes an inflammable gas. The vapour-and-air mixture thus formed enters the combustion chamber of the cylinder. A modern carburettor comprises four different systems: the main carburettor, the idling (slow-running) system, the acceleration pump, and the choke.

Combustion of the mixture in the cylinder requires oxygen; this is present in the air drawn into the cylinder when the piston descends. The air is passed through a filter (air cleaner) and a pipe called the induction pipe or intake manifold, in which the petrol is carried along with the stream of air (Fig. 1b). This effect is based on a law of physics known as Bernoulli's equation, which states that for a gas (or other fluid) flowing through a pipe the sum of the static pressure and the dynamic pressure is constant. This means that when the velocity, and therefore the dynamic pressure is increased, the static pressure decreases. If the induction pipe is narrowed to a reduced diameter at one particular section, the velocity of the air at that section will be increased, while the (static) pressure will diminish to a negative value in relation to the surroundings. In other words, a suction is developed there, which causes the petrol to be sucked out of the choke tube and be atomised (Fig. 1b). The tiny droplets are carried along into the cylinder by the air stream. The air intake pipe is preferably laid close along the exhaust pipe. The heat given off by the latter serves to preheat the intake air and helps to vaporize the petrol droplets. The main jet, which is essentially a constriction in the fuel supply pipe from the float chamber to the choke tube, limits and controls the quantity of fuel introduced into the air stream per unit of time: the jet has a small orifice which allows the fuel to flow from the float chamber only at a certain limited rate. The rate of supply of fuel to the carburettor itself is controlled by the equipment. This comprises the float chamber, the float and the needle valve. When the float chamber fills up with petrol, the hollow metal float rises until the needle is in contact with the valve seat and thus closes the petrol inlet opening. By this means the level of the petrol in the chamber can be constantly maintained at 2–3 mm (about $\frac{1}{8}$ inch) below the opening of the outlet pipe. When the float rises, the air which it displaces is forced back into the induction pipe. The throttle valve, which is operated by the accelerator pedal, is closed when the engine is idling. In this condition the main system is out of action and the pilot system comes into operation instead. The latter system provides the requisite combustion mixture for idling, i.e., when the engine is running under zero load (Figs. 3a and 3b), and operates independently of the main system. With the throttle closed, only a small amount of air can flow through the induction pipe and is unable to carry along any petrol from the choke tube. However, at a narrow gap between the throttle valve and the carburettor wall the air velocity is increased to such an extent that a powerful suction is developed behind the valve, causing petrol to be sucked from the pilot outlet. The necessary air is drawn through a passage branching off from the main air inlet at the top. Additional air is drawn through another branch passage, which is provided with an air regulating screw. By turning this screw in or out the composition of the pilot mixture (i.e., the fuel-and-air mixture for running the engine when idling) can be made leaner or richer (Fig. 3a). In another type of pilot system (Fig. 3b) it is not the quantity of additional intake air that is regulated; instead, the quantity of pilot mixture that is admitted through the pilot outlet into the induction pipe is varied by means of the mixture regulating screw. In this case, the upper outlet comes into operation when the throttle valve is slightly opened. The "idling" position of the throttle valve is determined by a stop.

(Continued)

1. Carburetor in U.S.A.
2. Gasoline in U.S.A.



ig. 1a MAIN CARBURETTOR SYSTEM NOT FUNCTIONING

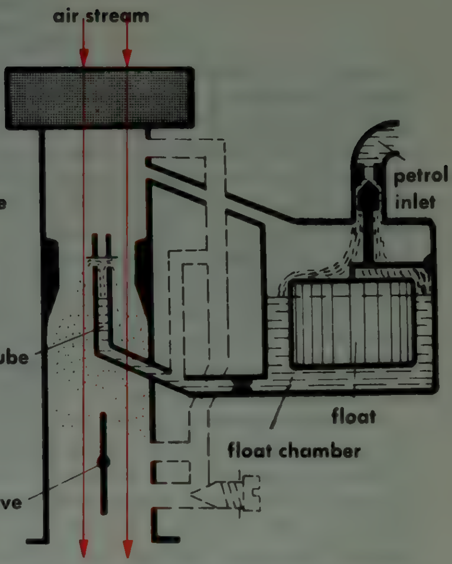
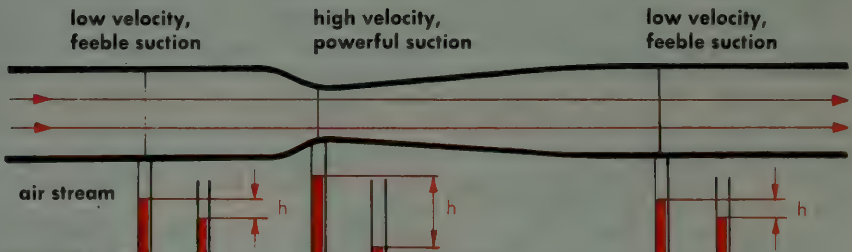
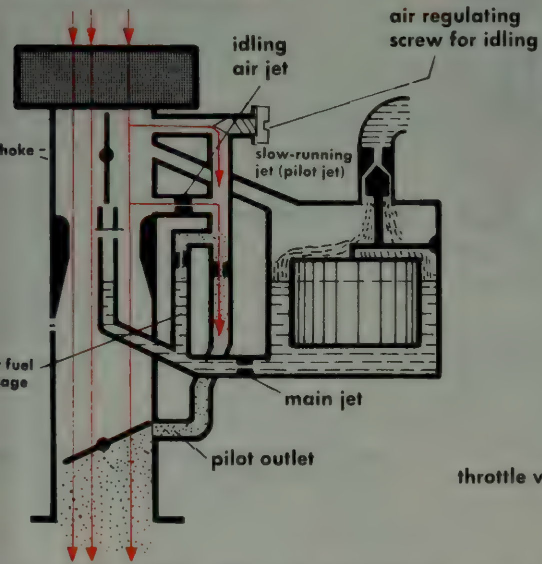


Fig. 1b MAIN CARBURETTOR SYSTEM FUNCTIONING



ig. 2 PRINCIPLE OF
HE INDUCTION PIPE



3a PILOT SYSTEM WITH AIR REGULATION

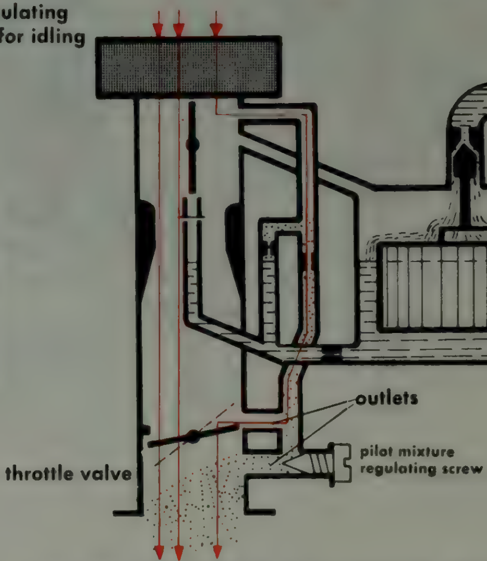


Fig. 3b PILOT SYSTEM WITH MIXTURE REGULATION

CARBURETTOR
(continued)

Sudden acceleration of the engine speed calls for an immediate increase in powe output, and this in turn requires a momentarily richer combustion mixture, i.e., mixture containing a higher proportion of petrol. If the throttle valve is suddenl opened when the engine is running at a low speed, the suction developed by the lo air velocity in the induction pipe will not be sufficient by itself to draw enough petr from the choke tube to raise the engine speed. This problem can be solved either b providing an extra supply of petrol in readiness in a storage chamber (Fig. 4a) t boost the normal flow when required, or alternatively by providing an injection pum (Fig. 4b) which is connected to the accelerator pedal. When this pedal is depresse the spring is slackened by the pump linkage, and petrol flows through the outle valve and the jet into the choke tube. When the pedal is released, the pump pisto rises, the outlet valve closes, more petrol enters through the inlet valve, and th spring is tensioned again.

When the engine is started from cold, a high proportion of the fuel in the fuel-and air mixture is precipitated on the cold wall surfaces of the induction pipe and cylinde In order nevertheless to obtain a combustible mixture in the combustion chambe the carburettor must temporarily supply a very rich mixture. To this end, the mai air inlet in the carburettor is closed by a valve called the choke (Fig. 5). The throttl valve is then only slightly opened. When the piston in the cylinder of the engine de scends, a powerful pumping action is developed in the carburettor, and the suctio draws an ample flow of petrol both from the choke tube and from the pilot system When the engine fires, a poppet valve (Fig. 6) enables a somewhat larger quantity o air to be drawn in, since an excessively rich mixture is not suitable for the engin as it warms up. The suction developed by the engine causes the plate to be lifte against the restraining force of the spring, so that the supply of air to the engine i increased. At the end of the warming-up period, the choke is opened either by hand o automatically. The automatic action is, as a rule, thermally controlled. A coil sprin keeps the choke closed. This spring is a bimetallic one, i.e., it consists of two metal which undergo different amounts of expansion on heating, causing the spring tensio to change. As the engine warms up, this coil spring is also warmed and opens th choke an appropriate amount.

Fig. 6 shows the construction of a down-draught carburettor. Names like down-draught, up-draught or cross-draught carburettor indicate the direction of flow of the intake air.

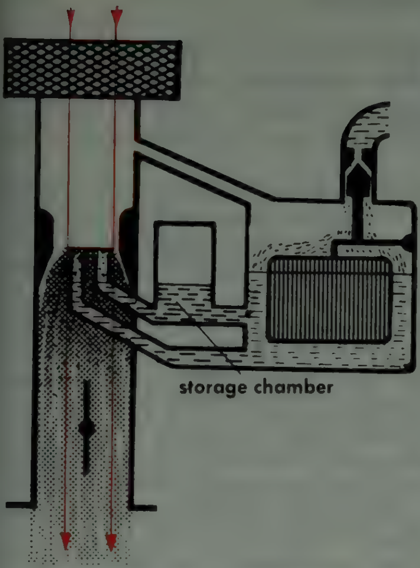


Fig. 4a ACCELERATION SYSTEM WITH STORAGE CHAMBER

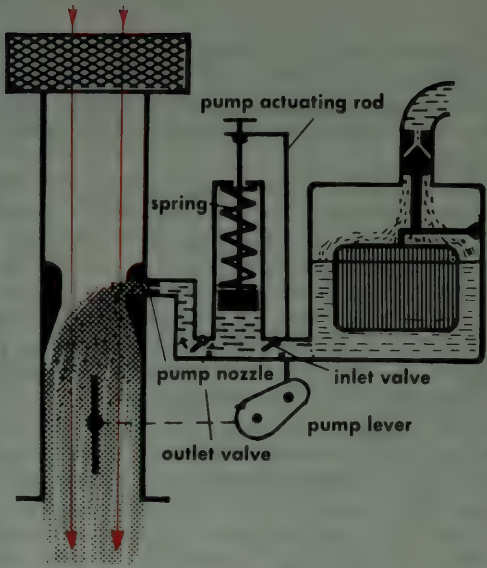


Fig. 4b ACCELERATION SYSTEM WITH INJECTION PUMP

Fig. 5 HOW THE MIXTURE IS FORMED WHEN STARTING FROM COLD

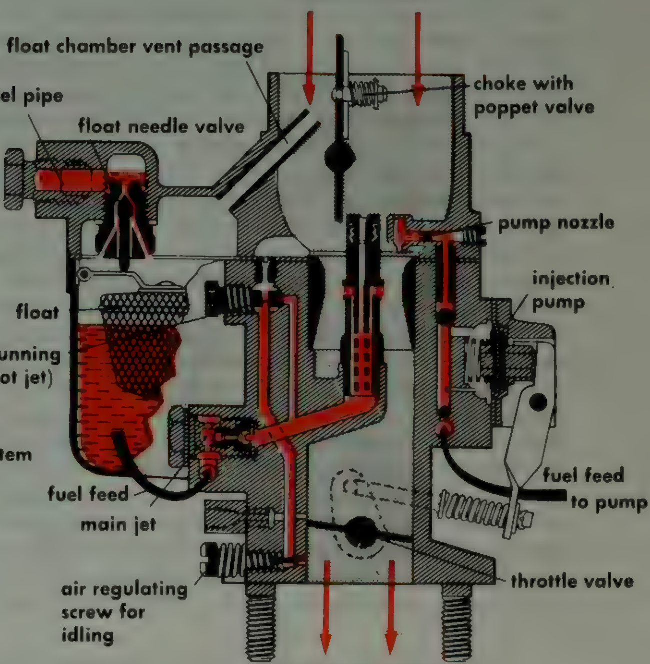


Fig. 6 SOLEX DOWN-DRAUGHT CARBURETTOR

The distributor is a device for conveying electrical current to the sparking plugs according to the firing order. It comprises the contact-breaker with condenser, the ignition cam, the actual distributor, and an automatic timing control device which determines the optimum ignition timing suited to the operating conditions of the engine.

The distributor has a longitudinal shaft whose lower end is connected to a drive pinion in the engine block. This shaft is rotated at the same speed as the camshaft of the engine and carries the ignition cam, which actuates the contact lever (Fig. 1). This lever is rotatably mounted on a plate. Fitted to the upper end of the shaft is the distributor rotor, which is made of plastic and through which an electrode passes. The distributor disc is also the cover of the housing. At the centre it is provided with a carbon brush through which the current is passed to the distributor rotor. At the edge of the disc are a number of tungsten electrodes, one for each cylinder of the engine. These are so arranged that the rotor is always at a contact just when the contact-breaker interrupts the circuit of the primary winding of the ignition coil. At that instant a high voltage is induced in the secondary winding of the ignition coil, and this voltage is allowed to pass through the rotor to the appropriate sparking plug. The condenser prevents the occurrence of sparking at the make-and-break contact. The ignition spark is produced as follows: so long as the make-and-break contact has not been opened by the cam, current can flow in the primary circuit (battery, primary winding, make-and-break contact, earth; Fig. 2), so that a magnetic field is formed in the ignition coil. At the instant when the contact interrupts the primary circuit, this magnetic field breaks down. This sudden change of the magnetic field induces a high voltage in the secondary winding, which voltage is thereupon applied to one of the sparking plugs, causing it to produce a spark.

The distributor housing also accommodates the timing control. This device automatically adjusts the optimum ignition timing, which very largely depends upon the type of petrol and upon the load and speed of the engine at any particular time (e.g., idling or running under full load). To adjust the ignition timing, the distributor shaft comprises two parts. The upper end is rotatable in relation to the lower end, so that the cam can be rotated relatively to the drive which is effected from below. A device operating with centrifugal weights rotates the upper end of the distributor shaft—therefore the cam—in relation to the drive by a certain amount corresponding to the speed of rotation of the engine. As a result of this automatic adjustment, ignition takes place earlier at high than at low engine speeds. When they are at rest, the centrifugal weights are pulled inwards by springs. When the speed increases, the weights are flung outwards by the centrifugal force until the latter is balanced by the spring force. A pin engaging with a slot in the centrifugal weights forms the connection with the top part of the distributor shaft. When the weights move outwards, the pin causes the upper end of the shaft to rotate in relation to the lower end; consequently, the cam is rotated relatively to the drive, and the instant at which the make-and-break contact opens is thereby altered. Additional adjustment can be controlled by the negative pressure in the induction pipe of the engine. This negative pressure, which develops behind the throttle valve in the induction pipe (page 202), cause a diaphragm to curve, and this movement is transmitted to the contact-breaker plate. As a result, the contact lever is shifted in relation to the cam, and the ignition timing is thereby altered. This adjustment operates more particularly when the throttle valve is not fully open (i.e., in the partial load range), whereas the ignition timing adjustment by the centrifugal weights is dependent upon the engine speed.

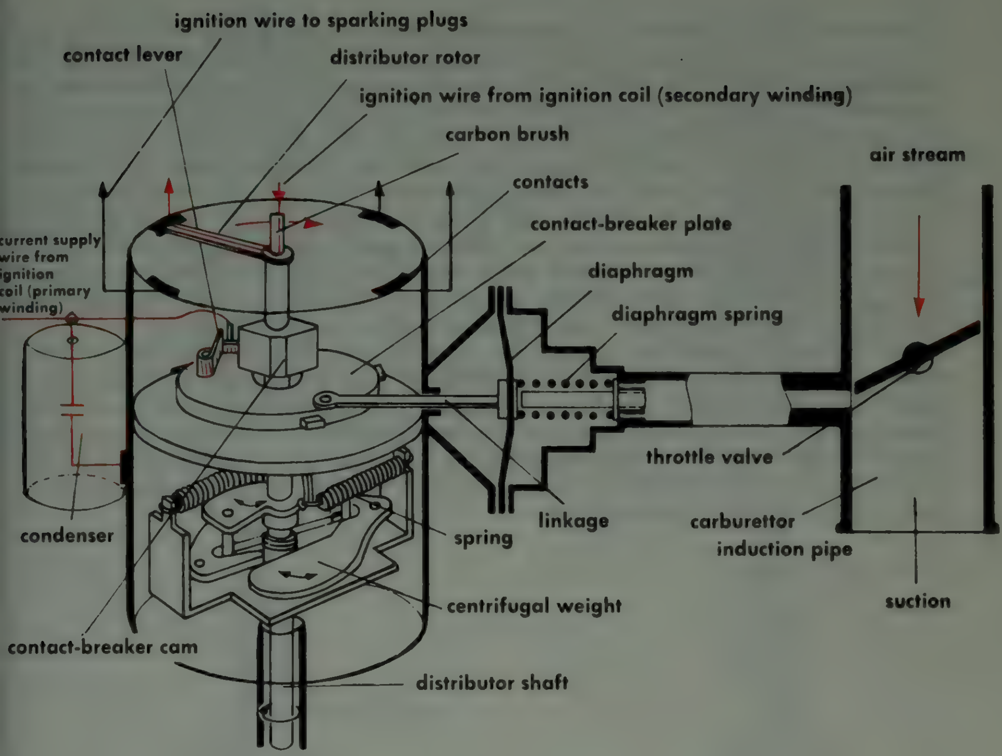


Fig. 1 DISTRIBUTOR WITH VACUUM-CONTROLLED SPARK ADJUSTMENT

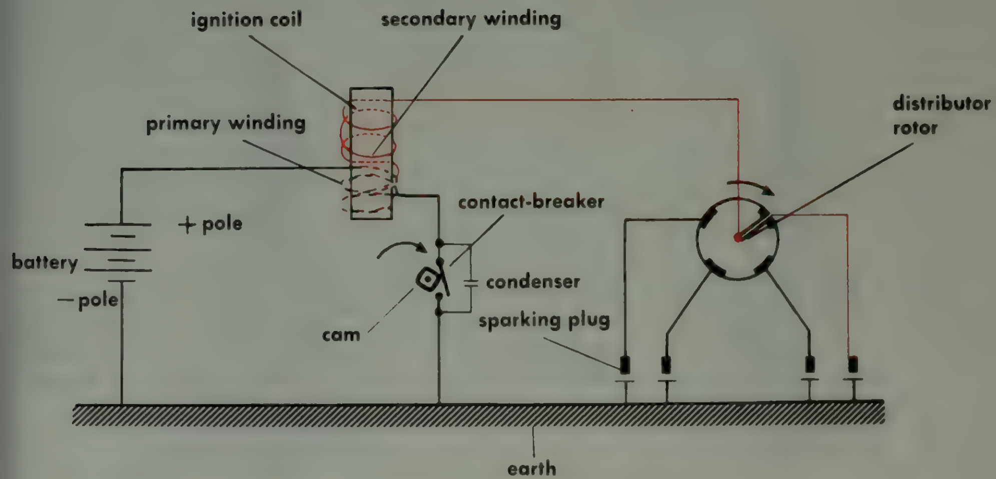


Fig. 2 DIAGRAM SHOWING OPERATION OF CONTACT-BREAKER AND DISTRIBUTOR

SPARKING PLUG[1]

The sparking plug ignites the compressed fuel-and-air mixture in the cylinder of a petrol engine by means of a spark which leaps from the central electrode to the so-called earth electrode. The sparking plug consists essentially of the two electrodes, the insulator and the body with its screw thread and connecting nut. The ignition current coming from the distributor (page 206) flows through the central electrode and produces the spark between this electrode and the earth electrode. The ignition voltage is about 25,000 volts, the spark gap between the electrodes being about 0.6–0.7 mm (0.025–0.03 in.). In engines with a low compression ratio, i.e., in which the combustion mixture is not so greatly compressed, the distance between the electrodes is made larger: this produces a longer spark, which can more readily ignite the mixture than a shorter one can. Since a highly compressed gas presents a higher resistance to the spark, a smaller spark gap is employed with high compression ratios.

The deciding factor in the choice of a sparking plug is its thermal value, which characterises the thermal behaviour of the plug. It is determined in a test engine under accurately defined conditions. For instance, a thermal value of 95 means that incandescent surface ignition occurs in the test engine after 95 seconds. A plug with a high thermal value, i.e., a plug which can be subjected to a high thermal load and yet remain cold, has a smaller insulator—so as to provide better heat dissipation—than a plug with a low thermal value.

1. Spark plug in U.S.A.

Fig. 1 MODERN SPARKING PLUG

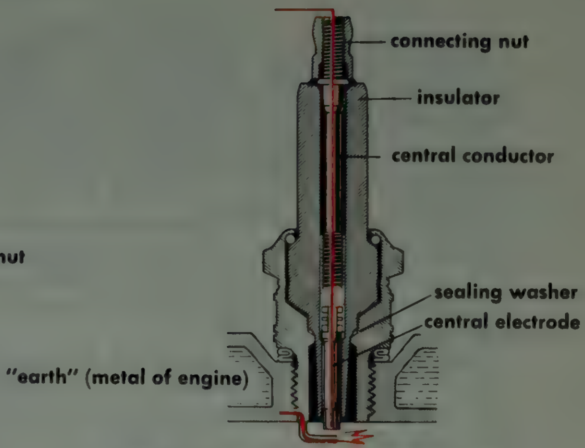


Fig. 2 SECTION THROUGH SPARKING PLUG (*functioning correctly*)

Fig. 3 INSULATOR DIRTY

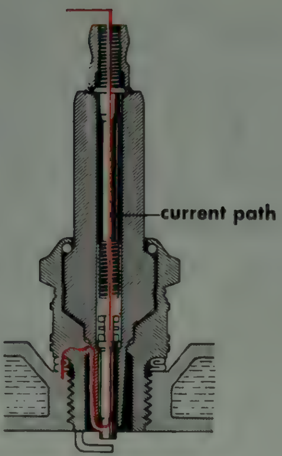


Fig. 4 TOP PART OF INSULATOR DIRTY

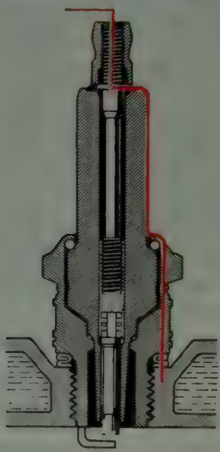


Fig. 5 SPARK GAP BRIDGED

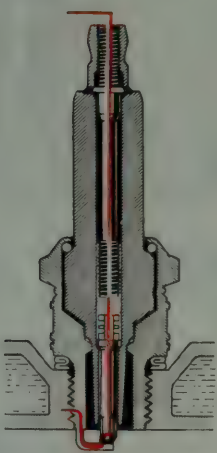
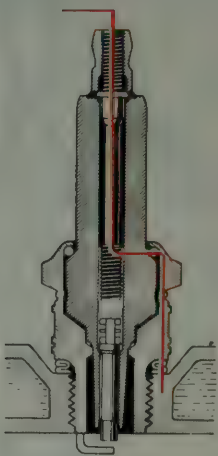


Fig. 6 INSULATOR FRACTURED



faulty plug operation (Figs. 3 to 6)

Internal combustion engines must have a certain minimum speed (approx. 300–600 r.p.m.) before they can run on their own power and develop a sufficiently high torque to drive the vehicle. While the vehicle is stationary, with its engine running, it is necessary to disconnect the engine from the gearbox; this is done by means of the clutch (Fig. 1). When the vehicle is about to move off, the clutch must first gradually raise the speed of rotation of the gearbox input shaft to the speed of the engine shaft. In so doing, the clutch has to transmit a torque while the slip due to the difference in speed between the engine shaft and the gearbox input shaft steadily decreases. Short periods of disconnection of the engine from the gearbox by means of the clutch are also necessary for gear-changing—i.e., for changing the transmission ratio of the gearbox—so as to effect the temporary disengagement of intermeshing components in the gearbox.

In the commonest form of clutch the connection of the engine shaft to gearbox shaft is effected by friction between two or more discs. The slip which occurs during the period of equalisation of the speeds of the driving shaft and the driven shaft generates heat. In the normal clutch operations the amount of heat involved is negligible, but if the clutch is allowed to slip for a considerable time, the clutch lining will quickly be destroyed.

Nowadays dry single-plate clutches are mostly used. In this form of construction a drive plate, made of steel and faced on both sides with riveted-on segmental clutch linings, is so mounted on the clutch shaft that it cannot rotate but can be shifted axially in relation to the latter (Fig. 2). The clutch shaft also forms the output shaft of the clutch to the gearbox. The drive plate is pressed between the flywheel and the clutch ring by thrust springs. This clutch ring is axially movable, but is so connected to the housing by drive elements (dogs) that it must always participate in the rotation of the housing. The clutch ring engages with the clutch release lever, the other end of which is moved by the movable clutch release sleeve. When the clutch is disengaged, the clutch release sleeve moves towards the engine, with the result that the clutch release lever shifts the clutch ring against the pressure of the thrust springs: the connection between the engine and gearbox is severed.

The multiplate clutch (Fig. 3) has been evolved from the single-plate clutch. Here again the clutch is integral with the flywheel of the engine. The housing contains the driving plates, which are non-rotatably connected to the housing, but can be shifted axially, and which are faced with clutch linings. The driven plates are secured to the pressure plate, upon which the pressure spring acts. The pressure plate is mounted directly on the drive shaft, on which it is axially movable, in order to enable the clutch to be disengaged (Fig. 3b).

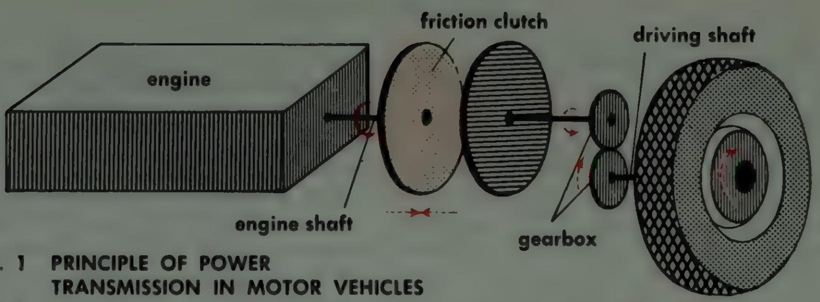


Fig. 1 PRINCIPLE OF POWER TRANSMISSION IN MOTOR VEHICLES

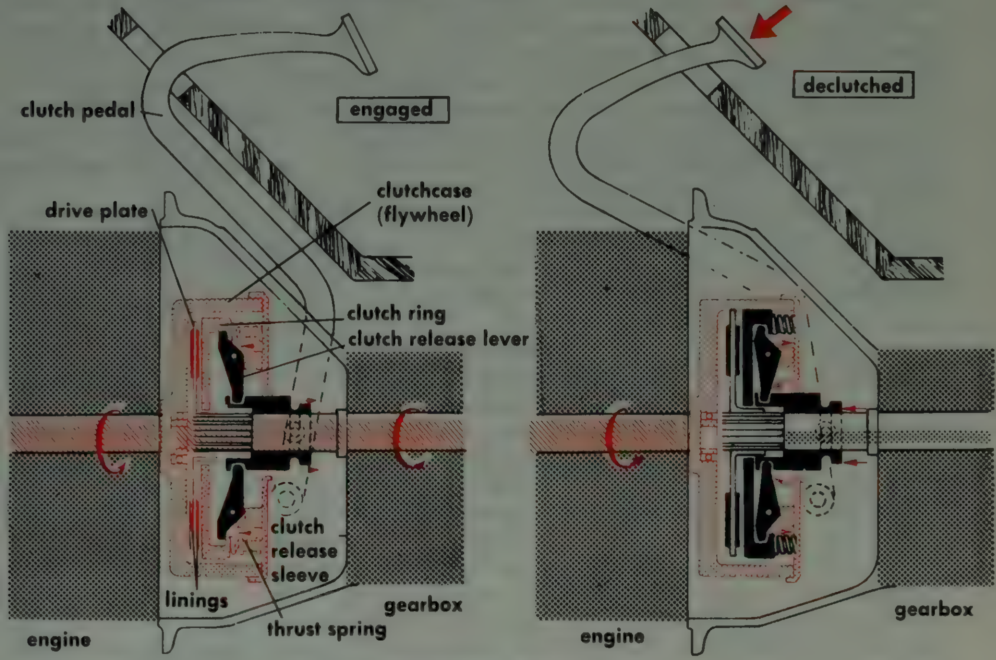


Fig. 2 SINGLE-PLATE DRY CLUTCH

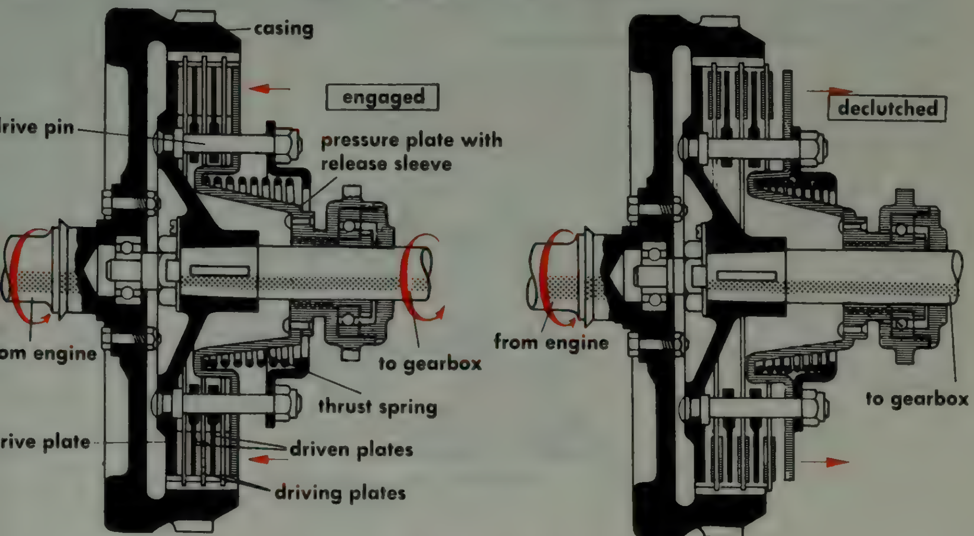


Fig. 3 MULTIPLATE DRY CLUTCH

An automatic clutch installed in a motor vehicle makes for much more comfortable driving in comparison with an ordinary pedal-operated clutch (see page 210). The clutch pedal inside the vehicle is obviated, since the clutch automatically disengages when the engine is idling and also when the motorist takes hold of the gear-shift lever. In principle, an automatic clutch comprises—besides the gear selecting clutch, which may be mechanically, pneumatically, hydraulically or electrically actuated—an additional starting clutch which may be engaged, for example, by centrifugal action or may be in the form of a fluid flywheel clutch or a magnetic powder clutch.

In the Saxomat automatic clutch the *starting clutch*, which in this case is of the centrifugal type, is a plate clutch which transmits power as soon as the centrifugal weights are flung outwards when the speed of the engine exceeds the idling speed. These weights press the pressure plate against the drive plate and thrust the latter against the flywheel (Fig. 1). At low speeds of rotation this clutch disengages automatically. In order nevertheless to take advantage of the braking action of the engine when travelling downhill, a freewheel is installed (Fig. 2), which establishes the mechanical connection between the crankshaft and the gearbox drive shaft as soon as the latter tends to rotate faster than the crankshaft.

The *gear selecting clutch* is likewise a plate clutch, which is always engaged during starting and which is actuated by means of a vacuum-controlled servo mechanism. When the motorist takes hold of the change-speed lever (Fig. 3), a contact is closed, with the result that an electromagnet is energised. It moves the valve to the right, so that the intake pipe of the engine is connected to the servo mechanism. Its diaphragm is now subjected on one side to a vacuum, and on the other side it is subjected to atmospheric pressure, with the result that it pulls the clutch actuating lever to the "disengaged" position. When the change-speed lever is released, the flow of current to the electromagnet is interrupted, the valve returns to its initial position and disconnects the servo system from the induction pipe. Atmospheric air now slowly flows into the vacuum chamber of the servo through a nozzle with a very fine orifice (which is so narrow that it does not disturb the sequence of events described above). As the pressure on both sides of the diaphragm has now been equalised, the latter returns the clutch actuating lever to the "engaged" position. Engaging the clutch is expedited by depressing the accelerator pedal, as the vacuum which is developed in the choke tube of the carburettor is applied, through a special pipe, to an auxiliary diaphragm, which lifts the valve "b", so that the atmospheric air can flow much more quickly into the servo. The tension of the spring which holds the valve "b" closed, and thus determines the vacuum pressure at which this valve opens, can be varied by means of an adjusting screw.

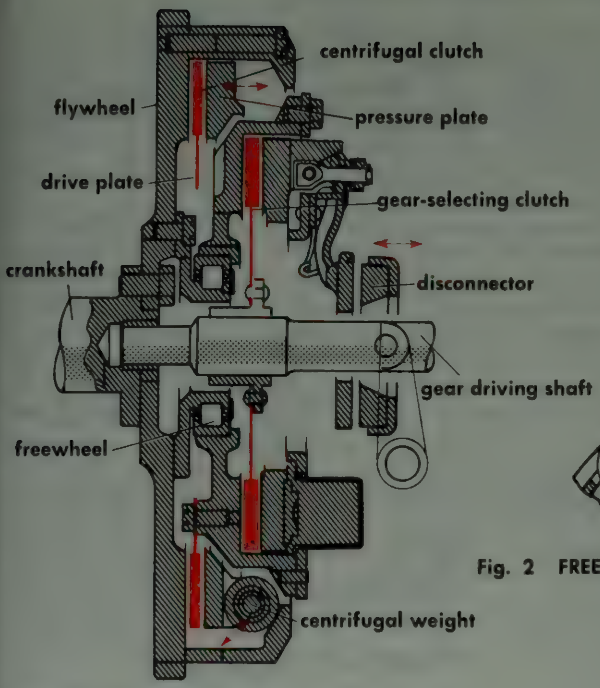


Fig. 1 SAXOMAT AUTOMATIC CLUTCH

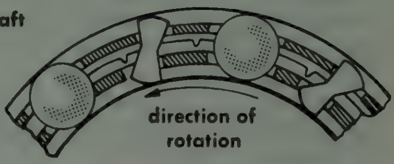


Fig. 2 FREEWHEEL WITH LOCKING MECHANISM

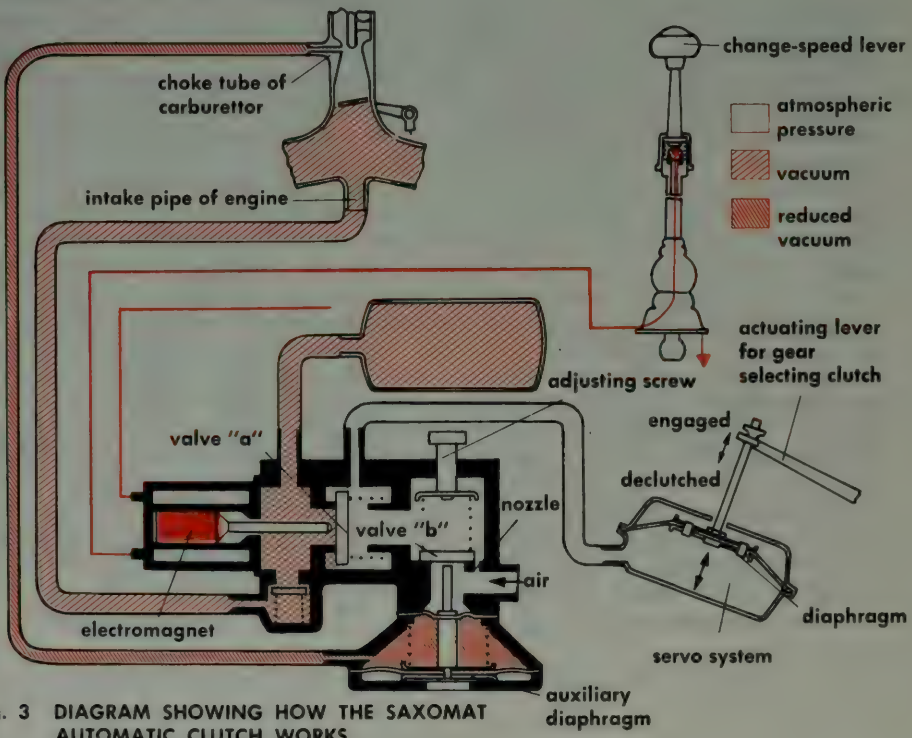


Fig. 3 DIAGRAM SHOWING HOW THE SAXOMAT AUTOMATIC CLUTCH WORKS

The internal combustion engines generally installed in motor vehicles develop their power output at high speeds of rotation (approx. 4000–6000 r.p.m.). It is therefore necessary to reduce the speed between the crankshaft and the shaft which drives the wheels. In addition, the torque (see below) that the engine delivers can be varied only within narrow limits. For this reason it is necessary to be able to alter the transmission ratio, so that the driving forces applied to the wheels of the vehicle can be adapted to the varying road resistance conditions.

Part of the requisite total speed reduction is effected in the final drive (see differential gear, page 224). The rest of the reduction can be obtained by means of the change-speed gearbox, which is interposed between the clutch (see page 210) and the final drive. In addition, the gearbox contains the mechanism for reversing. The gearbox alters the torque that is transmitted and therefore functions as a so-called torque converter. The transmission ratio (or gear ratio) is dependent upon the ratio of the numbers of teeth of the meshing gear wheels: a gear wheel mounted on a shaft rotating at a certain speed will drive another shaft whose gear wheel has half the number of teeth at twice the speed of rotation of the first shaft. Since the forces acting upon the two gear wheels are equal, the torque—i.e., the product of the force acting tangentially upon the gear wheel and the distance from the point of application of this force to the centre of rotation of the shaft—will, for the larger gear wheel, be double that for the small one. On the other hand, the smaller wheel will revolve at twice the speed of the larger. In other words, the smaller gear wheel has the higher speed, but the lower torque; the larger wheel rotates more slowly, but transmits the higher torque. The transmission ratio is the ratio of the input speed to the output speed; the pitch circle diameters of the gear wheels and torques of the shafts are in the inverse ratio.

A change-speed gearbox usually comprises the driving shaft end, the layshaft, and the driven shaft (Fig. 1), which are installed (parallel to the longitudinal axis of the vehicle) in the gearcase. A gear wheel is rigidly mounted on the driving shaft end which protrudes into the gearcase. This gear wheel is driven directly by the engine, through the clutch, and therefore rotates at the speed of the engine. It drives a second, somewhat larger gear wheel which is mounted on the layshaft, so that this shaft rotates at a lower speed. Rigidly mounted on the layshaft are the transmission gear wheels for the low speeds (1st, 2nd and 3rd gear in a four-speed, 1st and 2nd in a three-speed gearbox). The driven shaft—i.e., the shaft which transmits the desired speed to the final drive of the vehicle—is mounted in line with the driving shaft and carries the longitudinally movable driven gear wheels corresponding to the various speeds. The layshaft is rotating all the time. When the vehicle is in "1st gear" (Fig. 1), a small gear wheel on the layshaft drives a large gear wheel on the driven shaft; in "2nd gear" a larger gear wheel on the layshaft drives an only slightly larger gear wheel on the driven shaft (Fig. 2), so that the speed of rotation transmitted to the road wheels is somewhat higher than in "1st gear". In a four-speed gearbox a third pair of gear wheels on the layshaft and driven shaft drives the vehicle in "3rd gear". In "top gear" (direct drive) the engine speed is transmitted unreduced through the gearbox. For "reverse gear" the direction of rotation of the driven shaft is reversed by the interposition of a second gear wheel (Fig. 4). Many gearboxes are equipped with what is known as an overdrive or cruising gear: a large gear wheel on the layshaft drives a smaller gear wheel on the driven shaft, thereby causing the latter to rotate faster than the engine. This means that the engine need only run at a relatively low speed even when the vehicle is travelling fast.

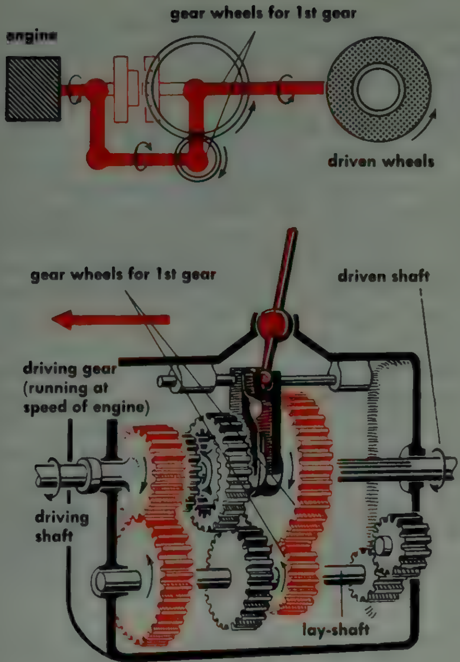


Fig. 1 GEARBOX WITH 1ST GEAR ENGAGED

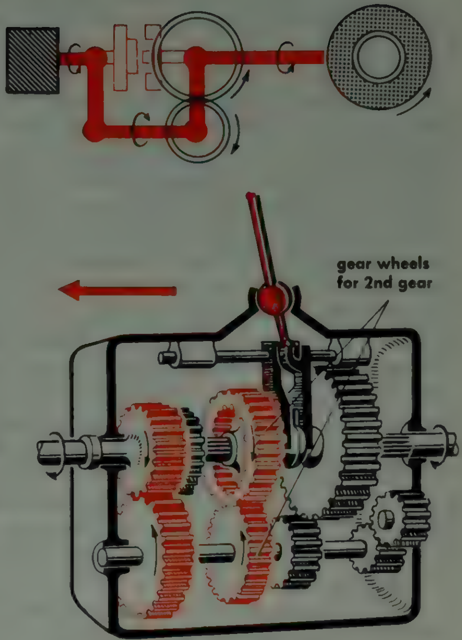


Fig. 2 2ND GEAR

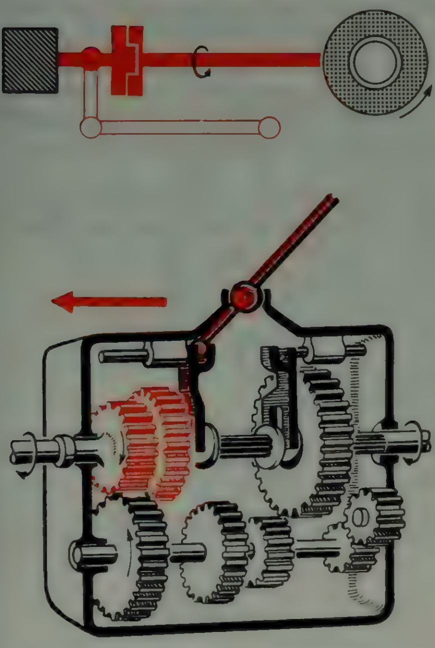


Fig. 3 3RD GEAR (DIRECT CONNECTION BETWEEN DRIVING SHAFT AND DRIVEN SHAFT)

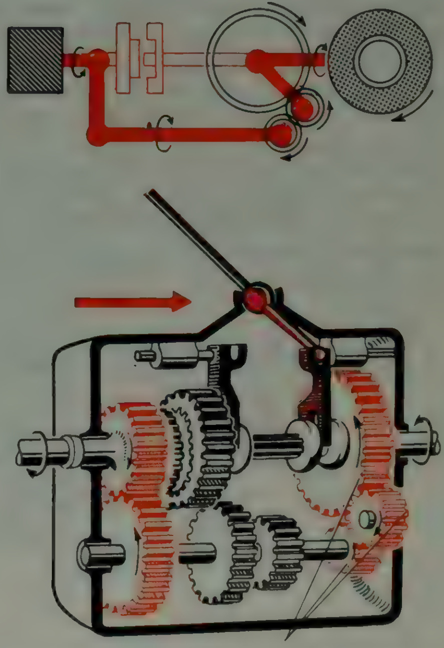


Fig. 4 REVERSE GEAR (DIRECTION OF ROTATION OF DRIVEN SHAFT IS REVERSED)

The fundamental difference between the conventional gearbox (see page 214) and the synchromesh (Figs. 1 and 4) is that in the former the gear wheels are brought into mesh—by sliding—only when the actual gear-change is performed, whereas in the latter all the pairs of gear wheels are constantly in mesh. The various transmission ratios are engaged by means of sleeves which are slid into position. Since the gear wheels themselves do not have to be moved in relation to one another, they can be provided with helical or spiral teeth for the sake of quiet running. In each pair of gear wheels one wheel is, for example, rigidly mounted on the layshaft, whereas the other wheel is loosely rotatable on the main shaft. To engage a particular ratio, the loose wheel of the pair of gear wheels concerned is locked to the shaft by means of dogs (locking elements). One set of dogs is mounted on the inside of a sleeve (the dog sleeve), and the other set is on the gear wheel which is to be engaged. The dog sleeve can slide axially on the main shaft but is locked to it in so far as rotation is concerned. However, before the dogs can engage with one another so as to transmit force, the dog sleeve (which is rotating at the same speed as the main shaft) and the gear wheel to be engaged (which rotates at a different speed) have to be synchronised, i.e., their speeds must become equal. This is done by means of small cone clutches or plate clutches. With cone clutches the gear wheel to be engaged is formed with a conical protrusion which slides into a conical socket in the dog sleeve. Plate clutches consist of small plates mounted on the shaft, which are pressed together by the pressure exerted by the sleeve, so that the two rotating parts, which are at first rotating at different speeds, are synchronised by slowing down and acceleration of the respective parts.

When a particular ratio (gear speed) is selected, the dog sleeve (Fig. 2) is shifted towards the gear wheel to be engaged (Fig. 3). The above-mentioned clutches thereupon come into action and synchronise the speeds of the two rotating parts. Then the dog sleeve can move farther forward and engage with the dogs on the gear wheel, just as if these two parts were stationary. The driving shaft is thus locked to the gear wheel. When engaging the selected gear, the motorist at first feels a resistance, which continues until the two rotating parts are synchronised. Only when this has been completed can the sliding sleeve be moved into its final position and the desired ratio thus definitely engaged. For the positions corresponding to the various ratios see Figs. 5–8 on page 219.

In automobile engineering a number of different kinds of synchromesh gear systems are used. They all function on the principle described here, but differ in various technical details.

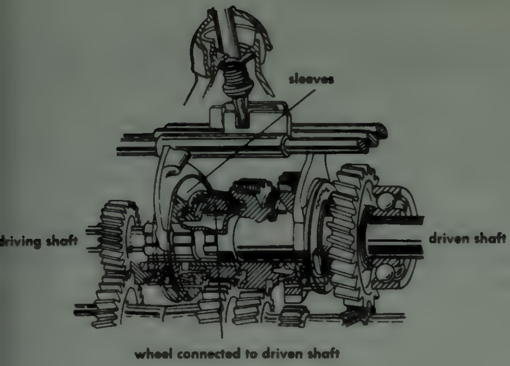


Fig. 1 FOUR-SPEED SYNCHROMESH GEARBOX
(3rd gear engaged)

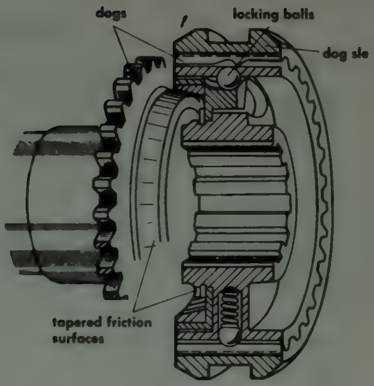


Fig. 2 SYNCHRONISATION *(cone clutch)*

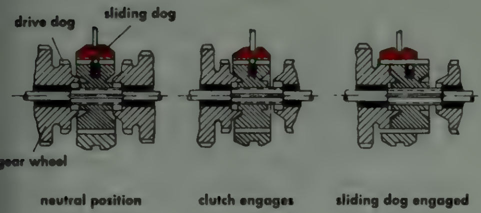
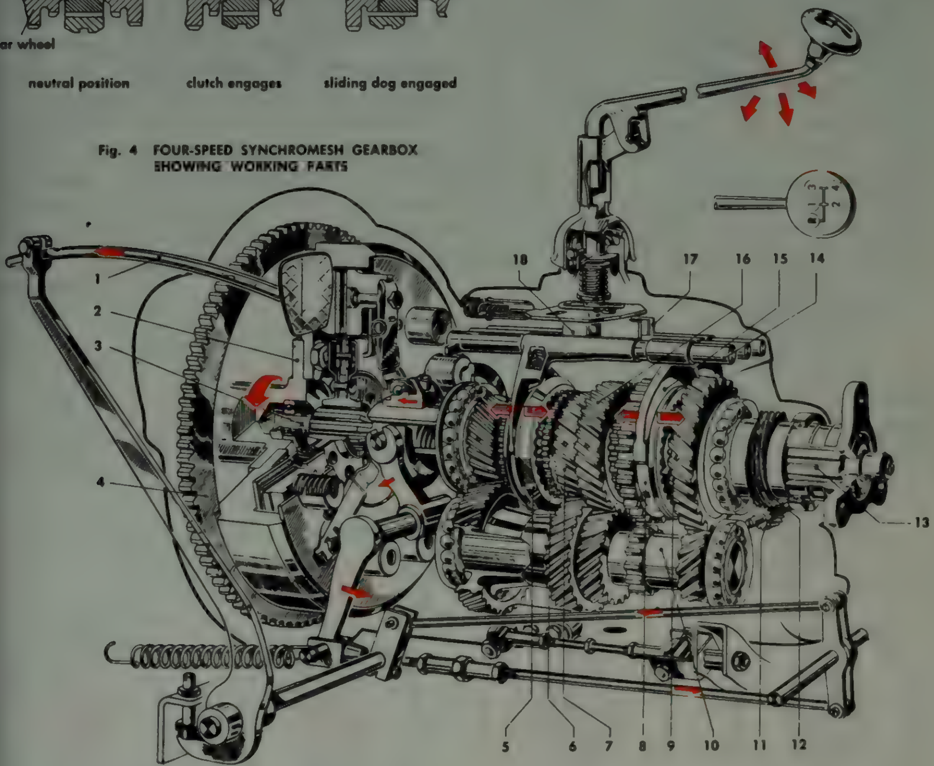


Fig. 3 HOW THE SYNCHROMESH WORKS

Fig. 4 FOUR-SPEED SYNCHROMESH GEARBOX SHOWING WORKING PARTS



1 clutch pedal
2 crankshaft
3 driving shaft
4 starting gear ring
5 sliding sleeve (3rd and 4th gear)
6 synchronising cone
7 spiral gear (3rd gear)
8 sliding sleeve (1st and 2nd gear)
9 spiral gear (1st gear)
10 lay-shaft
11 small drive wheel (speedometer drive)
12 spiral gear
13 main shaft
14 gear-shift rods
15 selector fork (1st and 2nd gear)
16 spiral gear (2nd gear)
17 selector head
18 selector fork (3rd and 4th gear)

In European automobile engineering, manual gear-changing is being superseded by automatic gear-changing in the larger cars. American private cars already nearly all have automatic transmission systems.

The automatic transmission has no gear lever for selecting the various speeds, or ratios. Speed control is entirely effected by the position of the accelerator pedal (gas pedal). Automatic gearboxes of this kind embody a combination of hydraulic and mechanical transmission components.

The hydraulic part comprises fluid couplings and torque converters; the mechanical part comprises planetary (or epicyclic) gears. The working principle of a fluid coupling can most conveniently be explained with reference to the behaviour of two fans placed together (Fig. 1): one of the fans is driven and produces a current of air which sets the blades of the other fan, which is not driven, in motion. The transmission of the rotary motion in this case is not effected by friction (as in the friction clutch; see page 210), but through the agency of a medium. In the example of the fans the medium is air; oil is used in a fluid coupling.

A coupling of this kind comprises two rotating parts fitted with vanes, one of which is the driving member (impeller) and the other is the driven member (turbine) (Fig. 2). The impeller itself is driven by the engine. The oil with which the coupling is filled is flung outward (Fig. 3) and, since it cannot escape, it is forced in between the vanes of the turbine, where it is deflected and flung back into the impeller. The circulating oil drives the turbine round at an increasing speed until it is rotating at the same speed as the impeller. A fluid coupling thus provides a smooth "take-up", so that the vehicle moves off entirely without any jerking. This transmits the same torque as that developed by the engine, except for minor losses due to so-called "slip", which is due to the fact that the turbine is not mechanically locked to the impeller but runs with a certain amount of lag in relation to it. At low speeds this loss is relatively far greater than at high speeds. Other important components of automatic gearboxes are the planetary gears (Fig. 4). They comprise a centrally mounted sun wheel and one or more planet wheels which engage with it and which furthermore engage with an external toothed ring (annulus). The possibilities presented by a gear system of this kind are indicated in Fig. 4 (cf. also page 184).

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synchromesh gearbox

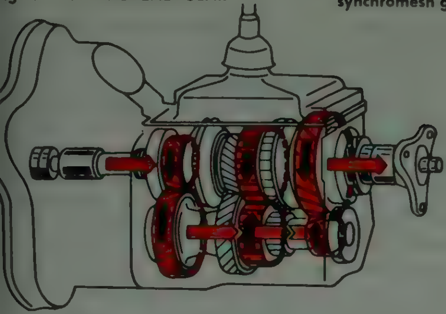

Fig. 5 1ST AND 2ND GEAR

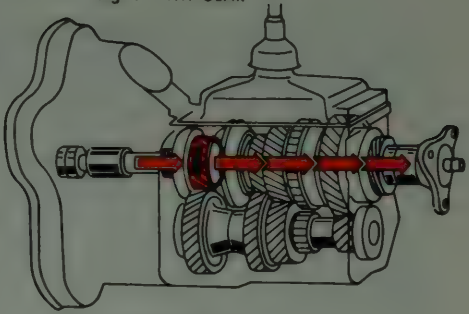

Fig. 7 4TH GEAR

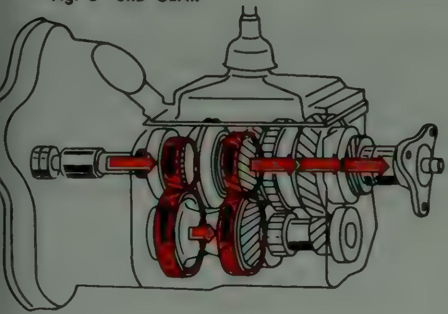

Fig. 6 3RD GEAR

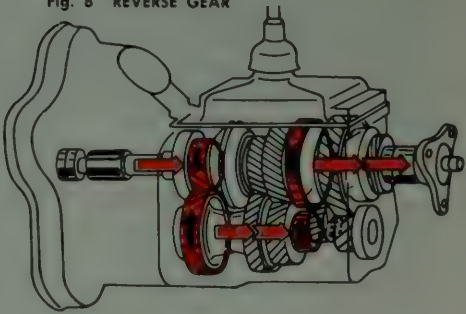

Fig. 8 REVERSE GEAR

automatic gearbox

Fig. 1 PRINCIPLE OF THE FLUID COUPLING

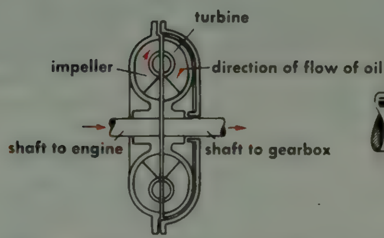

Fig. 2 FLUID COUPLING (*schematic*)

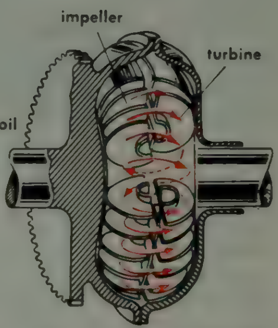

Fig. 3 HOW A FLUID COUPLING WORKS

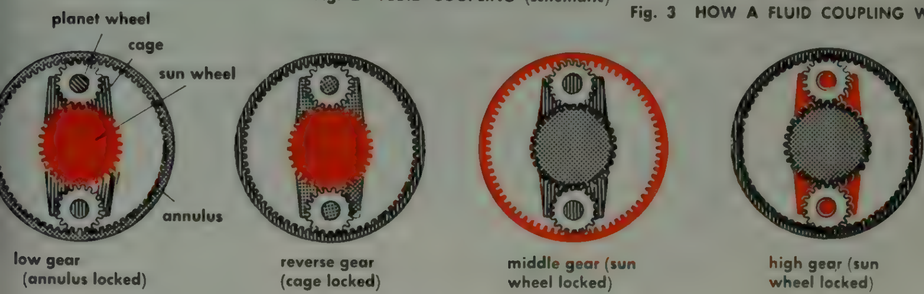

Fig. 4 PLANETARY GEARS

One of the most well-known forms of construction of an automatic transmission is the Hydra-Matic (Fig. 5; Fig. 7, page 223), great numbers of which were manufactured in the U.S.A. in the post-war years. It comprises two planetary gear sets for the four forward gears and one set for reverse. Transmission of force is effected by a fluid coupling. Gear-changing is effected automatically, depending on the speed of the car (speed of rotation of the driven shaft of the transmission), engine load, and accelerator position (through a governor). The transmission of force is effected in the mechanical part of the system by means of planetary gear sets and a fluid coupling. In first and second gear the total power is transmitted through the fluid coupling, which is driven at a higher speed in second gear than in first because the planetary gear set is then locked, and the slip losses in the coupling become less. On the other hand, the greater slip that occurs in first gear ensures great smoothness in take-up of the clutch.

In third and fourth gear, however, the power is transmitted in two ways. At the point marked *X* the power coming from the engine via the planetary gear set 1 is transmitted partly (through the fluid coupling) to the sun wheel (*S*) of the planetary gear set 2 and partly (through the hollow shaft and the locked clutch *K* 2) to the annulus (*A*) of the planetary gear set 2. In the last-mentioned gear set the two flows of transmitted power converge upon the planetary gear (*P*). The fluid coupling is by-passed. The slip then affects only that portion of the power which is transmitted by the coupling. In third gear the drive of the fluid coupling is effected through the operative planetary gear set 1, as in first gear, with reduced speed of rotation (1.44). In fourth gear, on the other hand, the planetary gear set 1 is locked by the clutch *K* 1. As a result of this the fluid coupling receives a higher operating speed in relation to third gear, so that the slip is reduced. In the parking position (*Pa*) the clutches *B*2 and *BR* are locked. *B*2 locks the annulus of *PL*2 and therefore the sun wheel of *PLR*. *BR* locks the annulus of *PLR*, so that the planetary gear between the sun wheel and the annulus is likewise immobilised. The drive to the rear axle is thus locked.

The gear-changing operations are controlled by a hydraulic regulating device. Fig. 6 shows the operations associated with changing up to the next higher gear. The centrifugal governor is driven by the driving shaft of the transmission. It comprises a centrifugal piston which is eccentrically and movably mounted in a housing. Controlled by the centrifugal force, whose magnitude depends upon the road speed of the vehicle, the centrifugal piston feeds a regulated supply of hydraulic pressure oil to the control valve. This oil pressure, regulated by the centrifugal governor, is now applied to one side of a piston, while a spring thrusts against the other side. This spring pressure is augmented by the oil pressure which is developed in the load regulator by the action of the accelerator pedal. If the oil pressure controlled by the load centrifugal governor exceeds the pressure controlled by the load regulator, the piston in the control valve clears the way to the relevant planetary gear set. At the same time the brake is released and the clutch is locked, as corresponds to the gear-changing operations described above. On the other hand, if the oil pressure controlled by the load regulator is greater than the pressure controlled by the centrifugal governor, the whole operation is reversed. The transmission system is thus controlled only by the accelerator (load regulator) in relation to the road speed (centrifugal governor). When there is no pressure in the pipe from the control valve to the relevant planetary gear set, then the spring action locks the brake and keeps the clutch released. A hydraulic regulator of this kind is provided for each planetary gear set.

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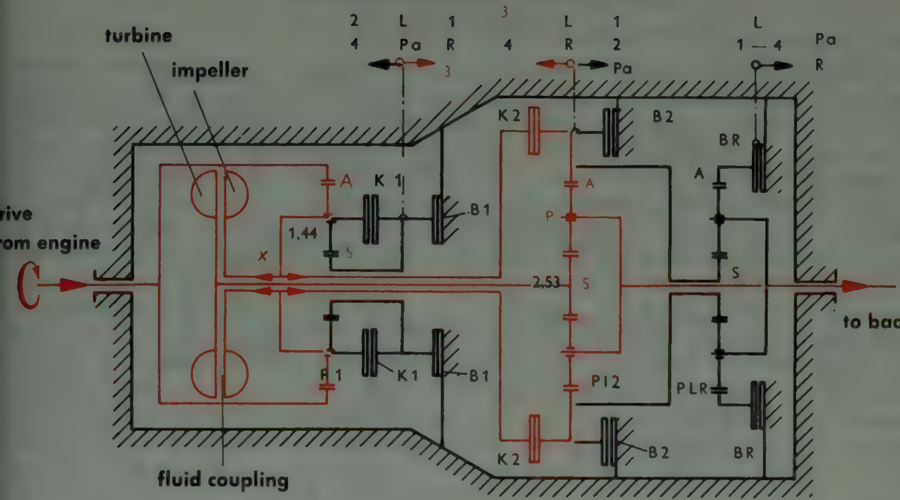


Fig. 5 DIAGRAM SHOWING HOW THE HYDRA-MATIC TRANSMISSION OPERATES (*3rd gear engaged*)

The various operating positions of this transmission are as follows:
Idling: All brakes (B) and clutches (K) released.

1st gear: Both brakes (B1, B2) locked, both clutches (K1, K2) disengaged; both planetary gear sets (Pl 1, Pl 2) are operative. Transmission ratio = 1.44 × 2.53 = 3.66:1.

2nd gear: Oil pressure from governor to Pl 1, so that B 1 is disengaged, K 1 locked. Only Pl 2 is operative; Pl 1 is locked by K 1. Transmission ratio 2.53:1.

3rd gear: Oil pressure to Pl 2, so that B 1 is locked, K 1 is disengaged, B 2 is disengaged, K 2 is locked. Only Pl 1 is operative. Transmission ratio 1.44:1.

4th gear (direct drive): Both brakes disengaged, both clutches locked; oil pressure to Pl 1 and Pl 2; Pl 1 and Pl 2 rotate in locked condition. Transmission ratio 1.0:1.

Reverse: B 1 locked, K 1 disengaged; B 2 and K 2 disengaged; BR locked (in the four forward gears it idles [L] and is always disengaged). Consequently PLR becomes operative, so that the direction of rotation of the driving shaft is reversed.

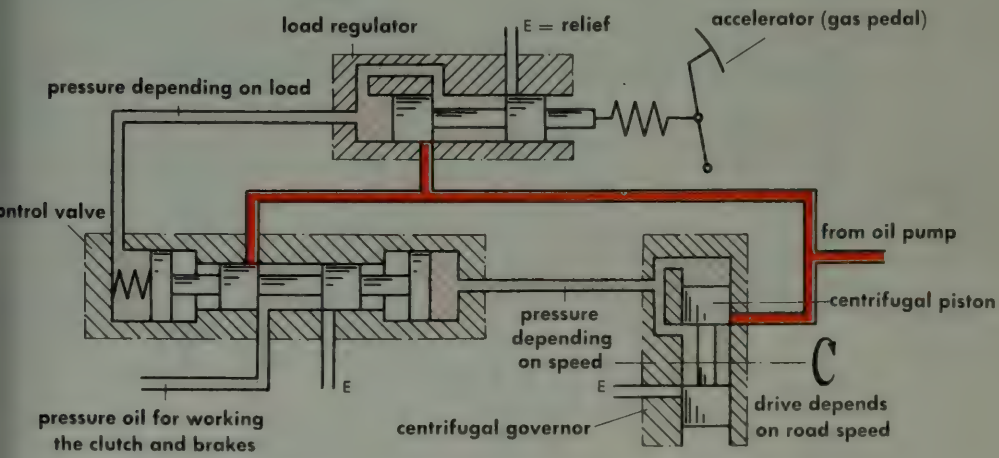
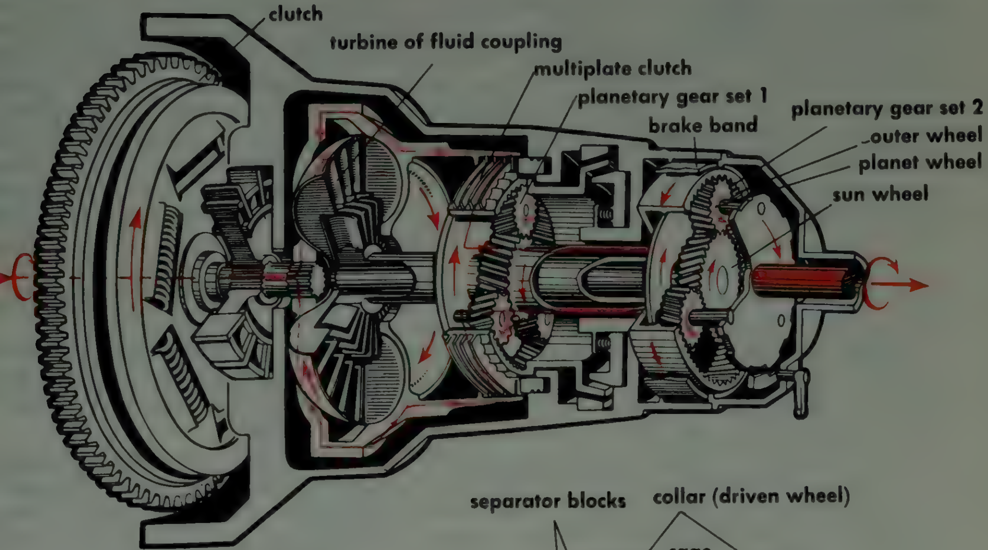


Fig. 6 HYDRAULIC GEAR-SHIFT SYSTEM OF THE HYDRA-MATIC

In other types of automatic transmission a fluid converter or hydraulic torque converter is used instead of a fluid coupling. Whereas the fluid coupling merely transmits the engine torque to the mechanical gearbox, the torque converter converts the low torque of the rapidly rotating engine into a high torque in conjunction with low speed of rotation at the output shaft of the converter. A torque converter of this kind therefore itself constitutes a transmission stage. In the fluid converter is an impeller which imparts its speed to the oil and forces it into the turbine. At low speeds of rotation the efficiency of such a device is very low, and for this reason the oil flowing out of the turbine is passed through a so-called stator (or reactor) whose fixed vanes are so curved that they redirect the oil flow into the impeller and thereby boost the action of the latter (Fig. 8). The stator vanes thus assist the conversion of the low torque of the rapidly rotating engine shaft into the high torque of the slowly rotating output shaft of the converter, so that the energy of the oil emerging from the turbine is not dissipated, but can be re-utilised. The speed of the turbine gradually becomes equal to that of the impeller, however, with the result that the efficiency diminishes again. Various arrangements are used to counteract this. In one of these the stator vanes are made adjustable, i.e., the deflection of the redirected flow is suited to the speed of rotation of the turbine at any particular time (in stationary systems). Another arrangement comprises a freewheel which thrusts against the casing during starting up. The force which restrains the freewheel is produced by the difference between the engine torque and the output torque of the transmission. When these torques finally become equal at the end of starting-up, the stator detaches itself from the housing, through the agency of a freewheel, and rotates along with the turbine. The converter has thus become a fluid coupling. The freewheel (Fig. 9) may be of the grip roller type. The transmission casing and the cage of the freewheel are rigidly interconnected. When the thrusting force of the stator acts in one direction, the grip rollers are jammed into narrowing gaps in the freewheel, so that the latter becomes locked and transmits force. When the thrusting force decreases or reverses its direction, the rollers are released, thus allowing the collar to rotate freely. The remaining gear ratios in an automatic transmission equipped with torque converters are obtained by means of planetary gear sets additionally provided (Fig. 10).

Fig. 7 SECTION THROUGH HYDRA-MATIC TRANSMISSION



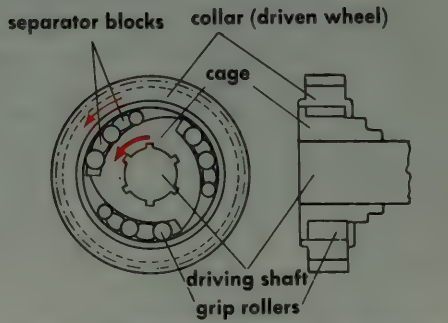


Fig. 9 FREEWHEEL WITH GRIP ROLLERS

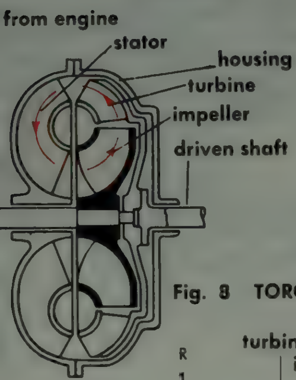


Fig. 8 TORQUE CONVERTER

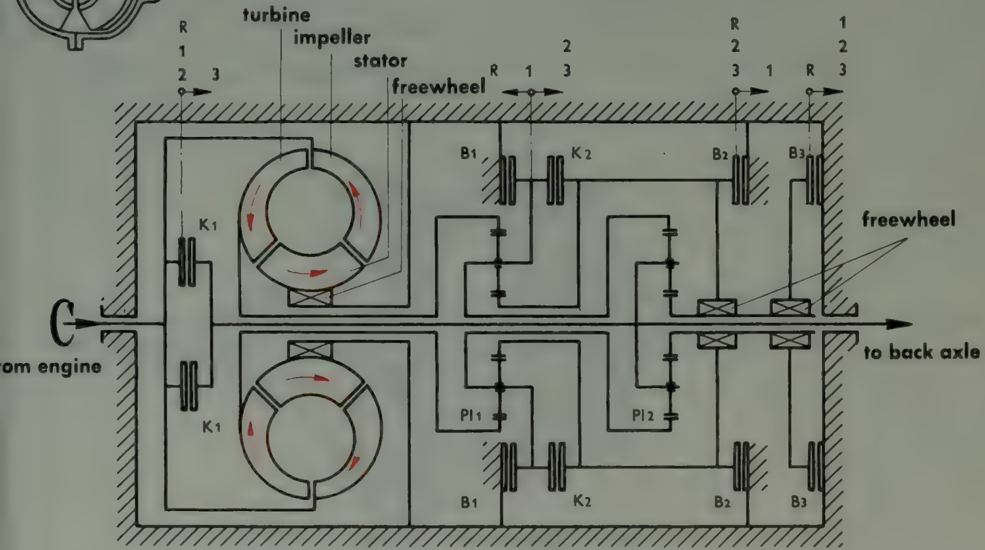


Fig. 10 BORG-WARNER AUTOMATIC TRANSMISSION WITH TORQUE CONVERTER AND FREEWHEEL

When a car turns a corner, the driving wheels (the driven road wheels) rotate at different speeds, since the inside wheel has to travel a much shorter distance than the outside one (Fig. 1). For this reason the power must not be passed straight to the two wheels mounted on one continuous axle, but must, instead, act through a device called a differential gear, which drives two separate half shafts—one to each driving wheel—with the same torque, but at different speeds of rotation. The propeller drive shaft, which forms the connection from the gearbox (or transmission case) to the rear axle, is provided at its rear end with a bevel pinion (Fig. 3), which drives a larger gear wheel called the crown wheel; the latter is loosely mounted on one of the half shafts of the rear axle but is firmiy connected to a box in which the four differential pinions are mounted. When this whole box rotates (as a result of the rotation of the crown wheel to which it is bolted), the differential pinions whose shafts are rotatably mounted in the box, will be carried round with it on their shafts (only two of the four differential pinions are shown in Fig. 3).

When the vehicle is travelling straight ahead (Fig. 3a), the box rotates, but the differential pinions will then not rotate on their shafts. They drive the bevel wheels which are rigidly connected to the inner ends of the driving shafts (the "half shafts" to the road wheels) and which now rotate at the same speed.

In Fig. 3b the vehicle is travelling in a bend: the right-hand half shaft is rotating at a slower speed than the left-hand half shaft. Now the differential pinions will rotate on their respective shafts, with the result that they retard the bevel wheel of the right-hand half shaft and at the same time accelerate the left-hand bevel wheel. The principle of this operation is shown in Fig. 2: the toothed racks symbolise the bevel wheels of the half shafts. If the two racks both move to the right at the same speed, the pinion will form a rigid connection between them. But if the bottom rack is retarded a little, the pinion will roll along it and, consequently, rotate in relation to both racks. The system as a whole will move in the direction represented by the arrow in Fig. 1.

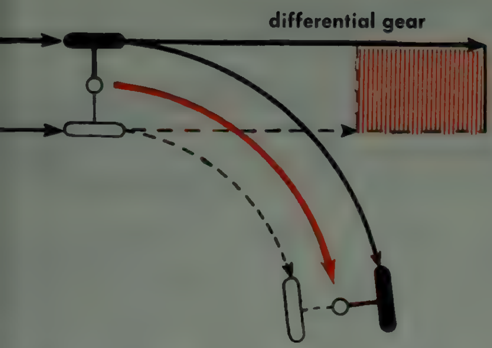


Fig. 1 DIFFERENCE IN DISTANCE TRAVELLED BY OFFSIDE AND NEARSIDE WHEELS IN CORNERING

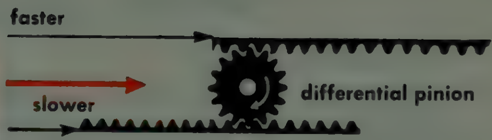


Fig. 2 PRINCIPLE OF DIFFERENTIAL GEAR

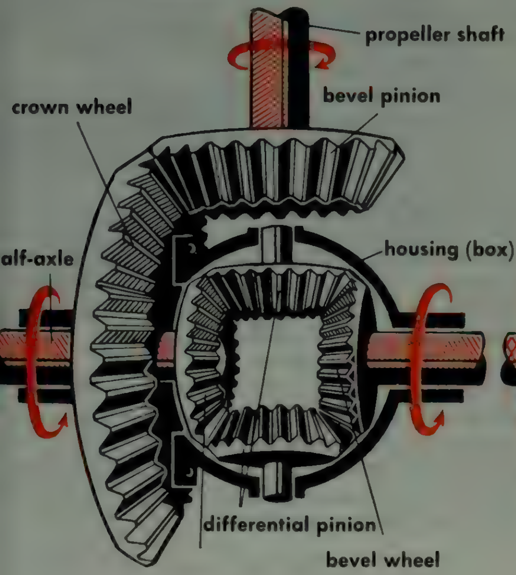


Fig. 3a DRIVING STRAIGHT AHEAD

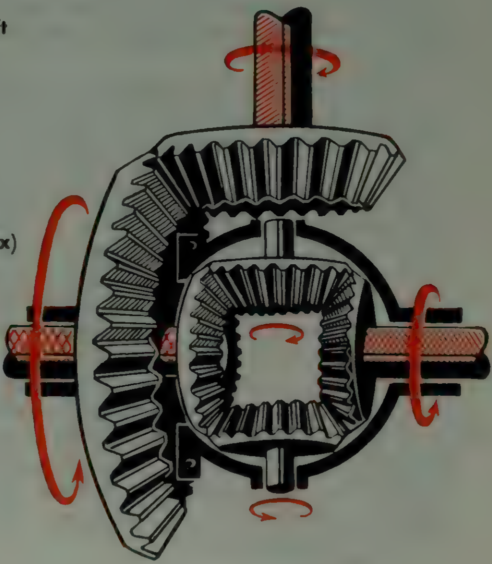


Fig. 3b CORNERING

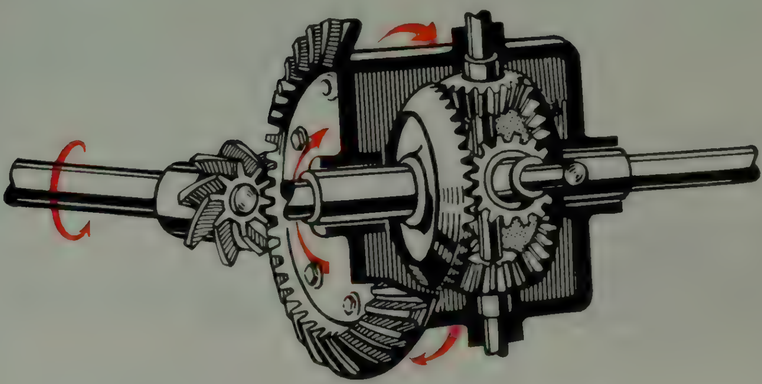


Fig. 4 DIFFERENTIAL GEAR

Springing

The object of the springing is to transform the short sharp jolts from the road into soft damped oscillations, so that only quite small forces are transmitted to the vehicle body itself. Also, the spring must prevent the wheels being lifted off the road as a sudden jolt occurs and thus safeguard the "grip" or "ground adhesion" of the vehicle.

Leaf springs are the oldest—and still frequently applied—springing system. Such springs consist of a number of steel strips (spring leaves) which are stacked one upon another and are held together by spring clips. The longest leaf of the spring is formed with an eye at each end through which a mounting pin is inserted (Fig. 1).

At one end the spring is mounted in a bearing bracket, while the other end is attached to a spring shackle which takes up the elongation of the spring when it deflects. Usually leaf springs of the semi-elliptic type are used as longitudinal springs with rigid axles (Fig. 2) or as transverse springs (Fig. 3). Longitudinal springs are fixed at their ends, the sprung mass being supported at the centre of the spring, whereas transverse springs are fixed at the centre, while the sprung mass is supported by the ends.

Coil springs are increasingly used in modern automobile engineering. They are lighter than leaf springs and require less maintenance, but they do not possess oscillation-damping properties (Fig. 4). What is desired of springs is that they should be very yielding at low loads and become stiffer at higher loads. A spring is said to be stiffer than another if it requires a greater force than the other spring to shorten or lengthen it a certain amount. In the case of a coil spring the force applied is proportional to the shortening or elongation it produces, i.e., it is equally stiff for all values of the load. Fig. 5 shows a progressively acting spring which fulfils the requirement as to increasing stiffness with increasing load. In this arrangement, a load increase first brings the main spring into action; next, the auxiliary spring and finally also the resilient rubber pads come into operation.

Fig. 6 shows the functioning principle of a *torsion bar suspension* system. One end of a torsion bar is fixed in the frame of the vehicle, while the other end is connected to a lever arm (pull rod). The spring actions are transmitted to the wheel by the lever (Fig. 7). They depend upon the thickness and length of the torsion bar; thick short bars are stiffer than thin long ones. Torsion rods consisting of bundled strips of spring steel may be used instead of solid rods of circular section.

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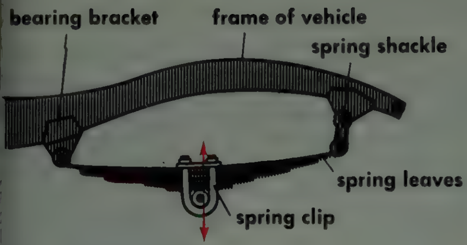

Fig. 1 LEAF SPRING

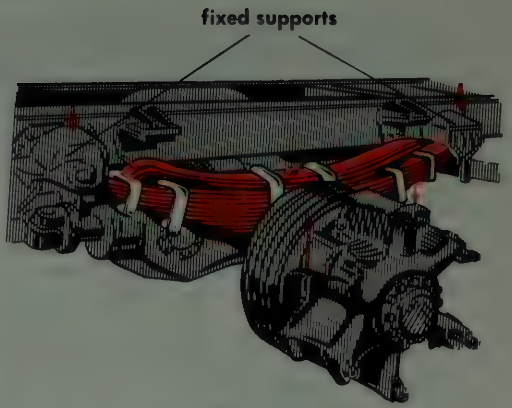

Fig. 2 LONGITUDINAL SPRING OF RIGID AXLE

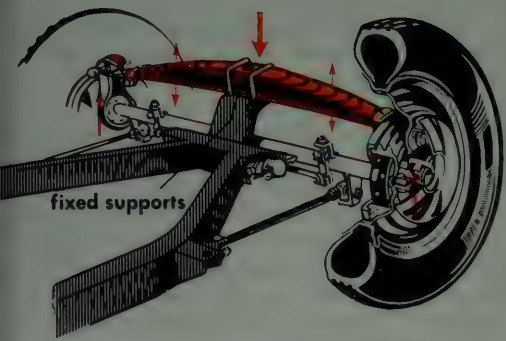

Fig. 3 TRANSVERSE SPRING
(*floating axle*)

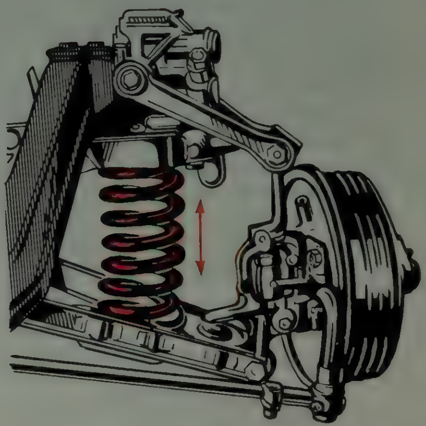
Fig. 4 COIL SPRING

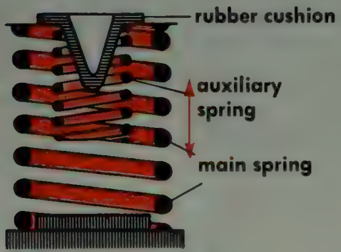

Fig. 5 COIL SPRING WITH PROGRESSIVELY ACTING AUXILIARY SPRING


Fig. 6 PRINCIPLE OF TORSION BAR SUSPENSION

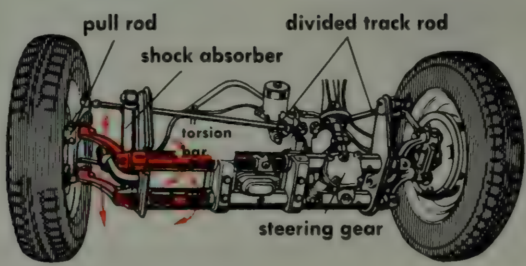

Fig. 7 FRONT AXLE WITH TORSION BAR SUSPENSION

With *air suspension* (pneumatic cushioning) the spring forces are resisted by a chamber filled with compressed air, which is formed either by a cylinder containing a piston which can move inside it or which may, alternatively, be designed as a compressible spring bellows. With this arrangement a suitably yielding suspension, well adjusted to the load, can be obtained, and the movement of the car body thereby reduced (more particularly when travelling in bends). As a result, riding comfort and safety are substantially increased. Depending on the type of suspension concerned, the air forming the cushion has a pressure of 3 to 9 atm. (45–130 lb./in.^{2}). The piston is connected directly to the sprung masses, which try to push the piston farther into the cylinder. This causes the air in the latter to become more highly compressed and thus resist this movement. The metal bellows (Fig. 8), which can absorb movements only in the direction of its longitudinal axis, is used mainly in conjunction with rigid axles. With this kind of suspension it is necessary to provide special guiding devices in order to obviate the occurrence of transverse forces (Fig. 9). Each bellows is connected to an auxiliary air vessel which ensures that there is always a sufficient quantity of air in the bellows. In the case of the roll bellows a piston-shaped movable body is directly connected to the auxiliary air vessel by the bellows (Fig. 10). It can also absorb transverse forces and can therefore be used in combination with independent wheel suspension and swinging half-axles (Fig. 11). The installation principle of air suspension systems is shown diagrammatically in Fig. 12. The road clearance is automatically always kept constant by means of the air equalising valve (height regulator). The latter is connected to the axle through a linkage system. When the load on the axle increases, the linkage rods will cause the piston of the height regulator to rise, so that air from the compressor and storage vessel can flow into the bellows, until the car body has risen so high that the inflowing air is cut off again by the piston in the height regulator. When the load on the axle decreases, or when the body lifts, the piston in the height regulator opens a hole through which air from the bellows can escape into the atmosphere, until the car body has regained its normal position and the piston in the regulator has returned to its middle position. The height regulator can be provided with a damping device, which produces a certain time lag in the response of the regulator. The regulating system will then not respond to jolts of short duration. In addition, damping prevents oscillation of the regulator about its normal position, i.e., continual alternation between admission of air into the pneumatic spring and escape of air.

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Fig. 8 AIR SUSPENSION
WITH METAL BELLOWS

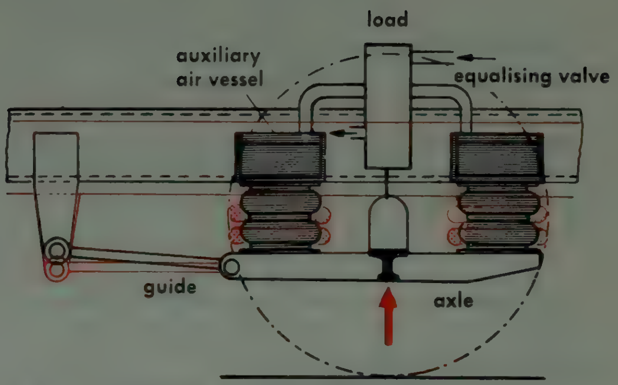
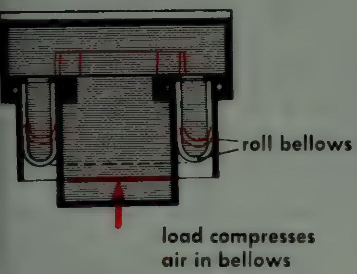


Fig. 9 EXAMPLE OF MOUNTING OF
METAL BELLOWS SUSPENSION



ig. 10 ROLL BELLOWS WITH INTERNAL
AND EXTERNAL GUIDANCE

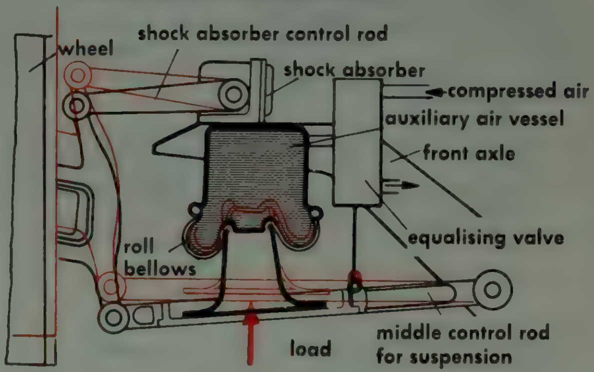
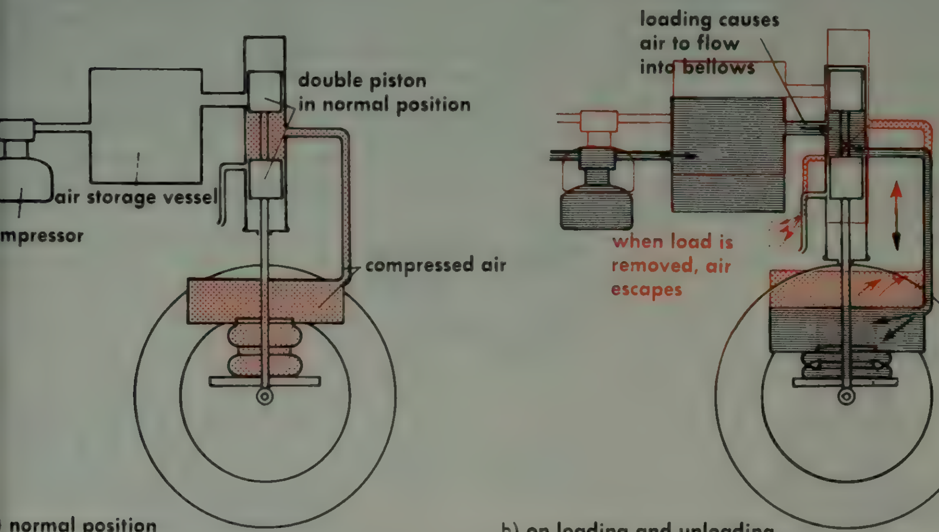


Fig. 11 EXAMPLE OF MOUNTING OF ROLL
BELLOWS SUSPENSION



normal position

b) on loading and unloading

Fig. 12 EQUALISING VALVE OF AIR SUSPENSION
SYSTEM (*height regulator*)

Axles and wheel suspension

The axles and wheel suspension determine the amount of the wheels in relation to the body of the vehicle. Various wheel movements are to be distinguished: (a) The wheel moves only in the vertical direction in its plane in relation to the body; (b) the wheel undergoes sideways displacement; (c) the angle of the wheel in relation to the body is changed; (d) sideways displacement occurs in combination with change of angle. Various forms of construction for the axles and wheel suspension are employed for eliminating these wheel movements as far as possible.

With a *rigid axle* (Fig. 14) the two wheels share the same suspension and springing. This type of axle is used chiefly for heavy vehicles, and more particularly for the rear axles. Such an axle often consists of a welded split sheet-steel casing. If one of the wheels encounters an irregularity on the road, this will cause the whole axle to tilt, with the result that the two springs are unequally stressed. The wheels undergo no parallel displacement when the axle tilts, but their angle does change. A *floating axle* (Fig. 15) is a rigid axle provided with a transverse spring above it. If the spring stop is located at the level of the centre of gravity of the vehicle, any tilting over of the body when the vehicle is negotiating a bend can be obviated. The floating axle therefore provides particularly good road-hugging ability during cornering (side-tilt stability).

The *swing axle* provides independent wheel suspension and springing, so that only the wheel that encounters a "bump" alters its position. For normal rear axle drive the independent suspension is usually provided in the form of the single-pivot swing axle (Fig. 16). The drawback of this arrangement is that the road contact surface of the wheel undergoes a certain amount of sideways displacement when the springs are compressed, for which reason it is not used for front axles (steering axles).

If it is desired to use the swing axle principle for the steering axle, the so-called wishbone suspension (Fig. 18) can be applied. This embodies a parallelogram system which ensures that the wheels are kept accurately parallel. However, here too, sideways displacement of the wheels occurs when the springs are compressed; in addition, angular displacement occurs when cornering. Another form of construction for steering axles is based on the use of double transverse springs (two leaf springs mounted one above the other: Fig. 17). The greater the deflections of these springs are, the larger will be the amounts of sideways and angular displacement of the wheels. The wheels may be mounted on arms, which are connected at one end to the frame and at the other end to the wheel axle. These arms may be single (Fig. 19) or double (Fig. 20). In this case, the springing takes the form of a torsion bar (see page 226). With suspension systems of this kind there is no sideways nor angular displacement when the springs deflect. The same is true of the arrangement illustrated in Fig. 21—in which the coil spring is enclosed in a tubular casing—except when the vehicle is cornering.

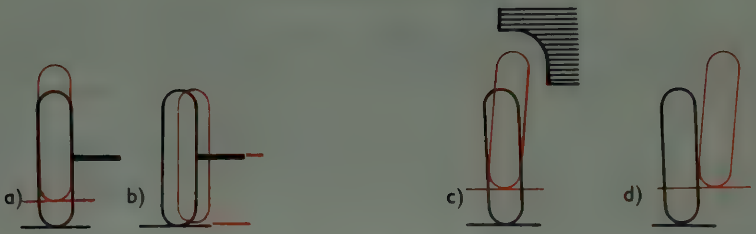


Fig. 13 WHEEL MOVEMENTS

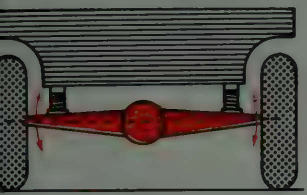


Fig. 14 RIGID AXLE

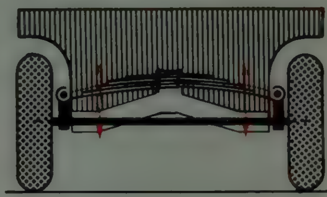


Fig. 15 FLOATING AXLE

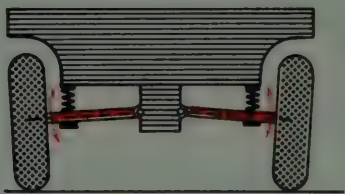


Fig. 16 SINGLE-PIVOT SWING AXLE

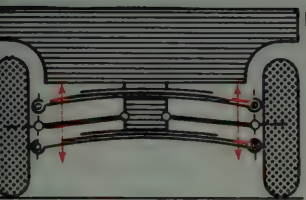


Fig. 17 DOUBLE TRANSVERSE SPRINGS

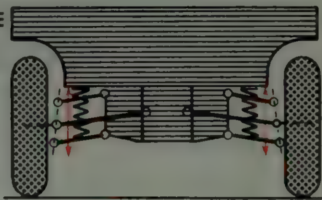


Fig. 18 WISHBONE SUSPENSION

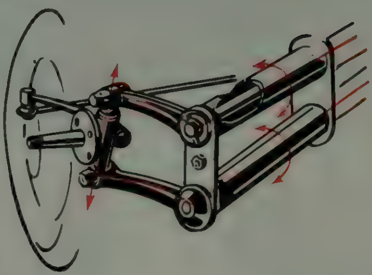


Fig. 20 TWO-ARM TORSION BAR SUSPENSION

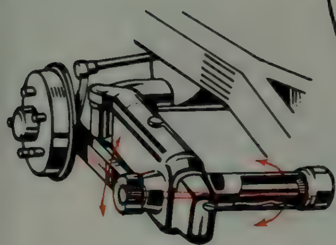


Fig. 19 SINGLE-ARM TORSION BAR SUSPENSION

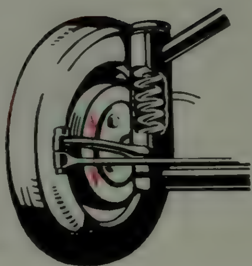


Fig. 21 INDEPENDENT WHEEL SUSPENSION WITH COIL SPRING IN TUBULAR CASING

SHOCK ABSORBERS

Shock absorbers are devices which are installed between the body of a motor vehicle and the wheel suspension and which perform the function of damping the objectionable spring oscillations that may be caused by irregularities of the road surface (Fig. 1).

In modern cars the shock absorbers are almost invariably of the hydraulic type: a piston moves inside a cylinder and forces oil through narrow holes or valves. In this way a high resistance to the movement of the piston is developed, so that it is greatly retarded (Fig. 2).

In the older types of shock absorber the piston travels to and fro in an oil-filled casing (Fig. 3). The movements caused by the road irregularities are transmitted to the piston through a lever system connected to the axle of the vehicle. When the lever moves upward, the piston goes to the left and forces oil through a passage and a valve to the other side of the piston. This upward movement must be only slightly damped, in order not to impair the spring action (Fig. 3a). A valve with a large orifice is accordingly used. This allows the oil to escape quickly, so that only a small damping effect is obtained. On the other hand, the return motion of the springing must be powerfully damped; for this reason a much narrower orifice is used for the other valve, which opens when the piston travels to the right and the lever moves downward (Fig. 3b). This shock absorber is therefore double-acting, but with different intensities of damping in the two directions. The type of shock absorber most frequently used nowadays is the telescopic shock absorber (Fig. 4). Its operating principle is the same as that of the lever type. It comprises two tubes, one fitting inside the other, the piston rod being connected to the outer tube; the piston moves in the oil-filled inner chamber of the inner tube. The piston contains so-called flap valves which alternately allow oil to pass in one direction only. They constitute the throttle valves that produce the damping action.

When the piston travels to the right (compression of the springs), the oil is forced through a flap valve or through holes into the left-hand chamber; when the piston moves back to the left (decompression of the springs), the oil flows through another valve back into the right-hand chamber. Since the left-hand chamber is smaller than the right-hand chamber (due to the presence of the piston rod in the former), some of the oil must, when compression of the springs occurs, flow through the bottom valve into the storage space. The bottom valve (Fig. 5) is so designed that it allows oil to pass in both directions, except that the flow resistance it presents when the springs are compressed is much higher because in that case the oil is not discharged through the orifice of the valve but is forced through the narrow gaps of the plates. On decompression of the springs, the oil which was forced into the storage space flows back into the inner chamber through the relatively large valve orifice, i.e., without encountering much resistance.

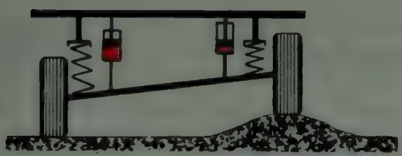
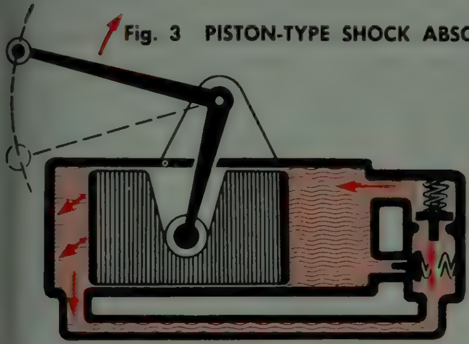


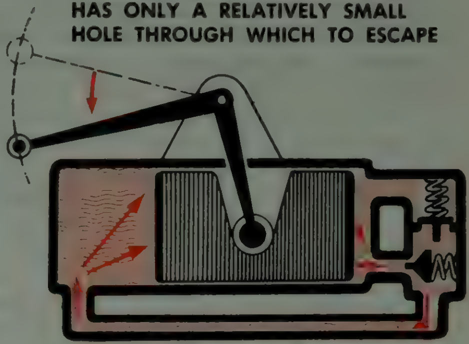
Fig. 1 TELESCOPIC SHOCK ABSORBERS

Fig. 2 THE PISTON MOVEMENT IS HARDLY RESISTED BY GAS, BUT A LIQUID OFFERS A HIGH RESISTANCE IF IT HAS ONLY A RELATIVELY SMALL HOLE THROUGH WHICH TO ESCAPE

Fig. 3 PISTON-TYPE SHOCK ABSORBER

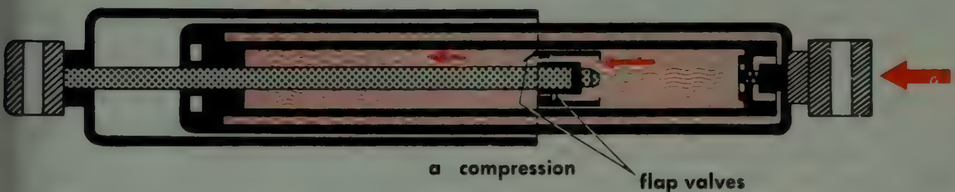


a slight damping

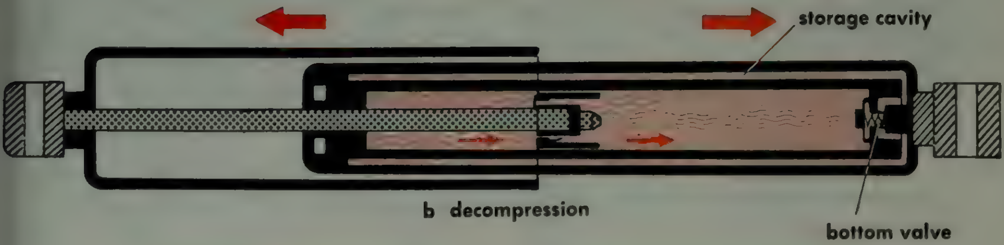


b powerful damping

Fig. 4 TELESCOPIC SHOCK ABSORBER



a compression



b decompression

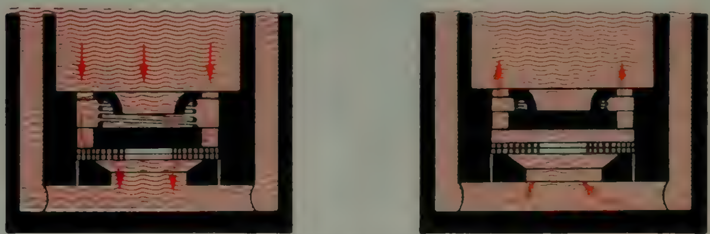


Fig. 5 BOTTOM VALVE POSITIONS IN COMPRESSION AND DECOMPRESSION

The function of the steering system is to enable the vehicle to change its directior of travel and negotiate corners. The main requirement applicable to a good steering system is to ensure geometrically precise rolling of the wheels (without slip) wher travelling in bends. In the case of motor cycles and three-wheelers with single-wheel steering this condition is automatically satisfied, because with such vehicles the centre of the circular motion is always located at the intersection of the extensions of the steered and the non-steered axle. The so-called bogie or fifth-wheel steering system (Fig. 1a) is geometrically satisfactory for four-wheeled vehicles. In this sytem the whole steering control shaft forms a rigid system which is rotated about its centre; it is usually employed on horse-drawn vehicles and trailers. The drawback of this steering system is that the stability of the vehicle decreases as the front axl and wheels swing farther round, and a further disadvantage is that the swing-roun of the wheels takes up a good deal of space under the frame. These draw-backs d not occur in the king-pin steering system (Fig. 1b), since each steered wheel has it own pivot point; on the other hand, it only approximately satisfies the main condi tion as to slip-free geometrically accurate rolling of all the wheels on the road surface As appears from Fig. 1b, the extension of the rear axle coincides at one point with th extensions of the two front stub axles only if the wheel on the inside of the curv has a larger angle of lock than the wheel on the outside of the curve, because onl then will all the wheels be rolling on concentric circles around the centre of the circula motion. The track arms (Fig. 1b) can, however, be installed, not parallel to the longi tudinal axis of the vehicle, but at an angle to it (so that the front axle, track arms an track rod form a trapezium—as in Fig. 1b—instead of a parallelogram). Then al the wheels will run approximately at right angles to the lines of connection to th common centre of the curved path.

In the king-pin steering system each wheel has its own pivot point, namely, th king pin, to which is attached the stub axle on which the wheel rotates. Each stu axle is connected to a short lever called the track arm. The two track arms are inter connected by the track rod. Thanks to this arrangement, the force applied by th steering control arm need act only at one king pin; the movements of the other kin pin will be automatically controlled by the trapezium linkage system. When th driver turns the steering wheel, the steering gear transmits the rotation of the whee to the drop arm and drag link which in turn actuates the above-mentioned steerin control arm. The latter then rotates the stub axle about the king pin (Fig. 2).

Figs. 3, 4 and 5 illustrated various forms of steering gear construction. In the worm and-sector system the bottom part of the steering column is provided with a worr which engages with a toothed worm-gear sector (Fig. 3). In the case of worm-and nut steering a nut moves along the worm when the latter rotates (Fig. 4). When th steering column is rotated, the worm-gear sector (Fig. 3) or the nut (Fig. 4) move up or down and transmits its movement to the drop arm.

A variant of the worm-and-nut system is the cam steering system (Fig. 5). In thi arrangement a tapered projection (follower) engages with the worm-shaped cam an moves the drop arm.

Fig. 1a BOGIE STEERING SYSTEM

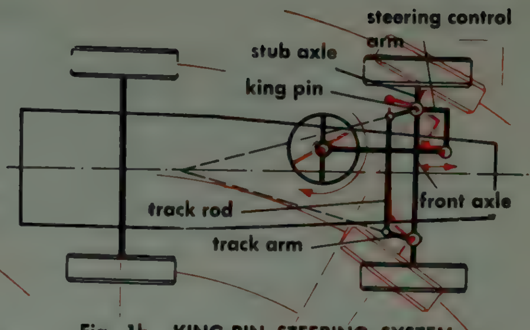


Fig. 1b KING-PIN STEERING SYSTEM

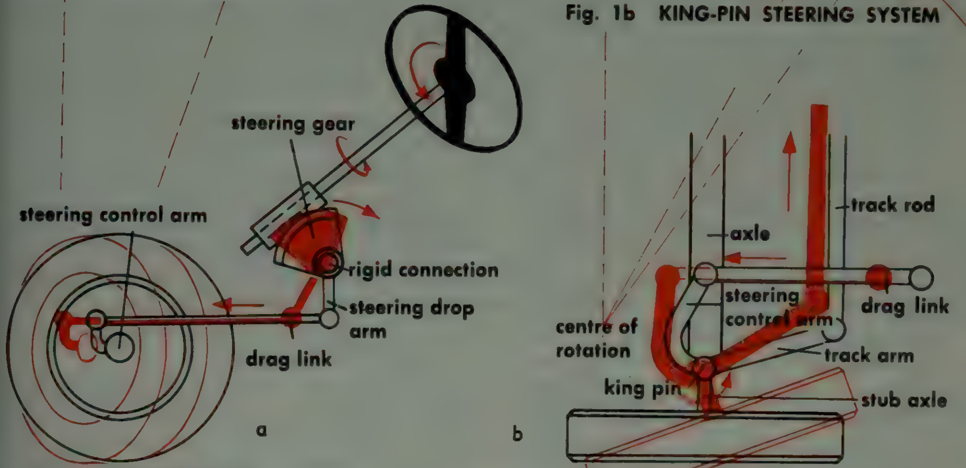


Fig. 2 DIAGRAM OF STEERING SYSTEM

a) side view b) top view

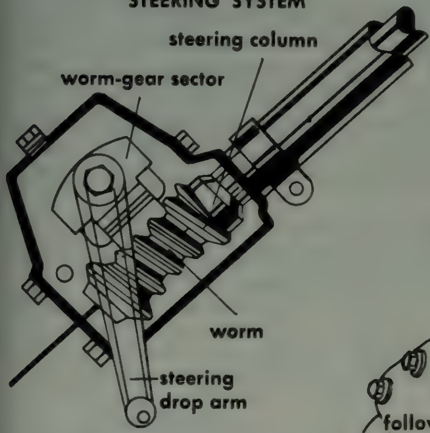


Fig. 3 WORM-AND-SECTOR STEERING SYSTEM

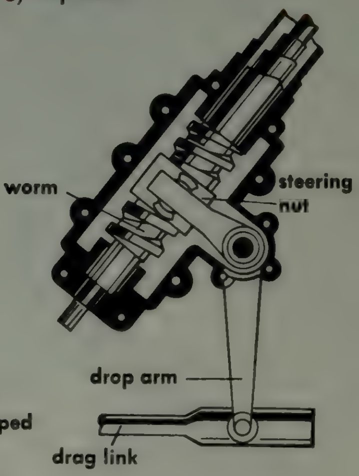


Fig. 4 WORM-AND-NUT STEERING SYSTEM

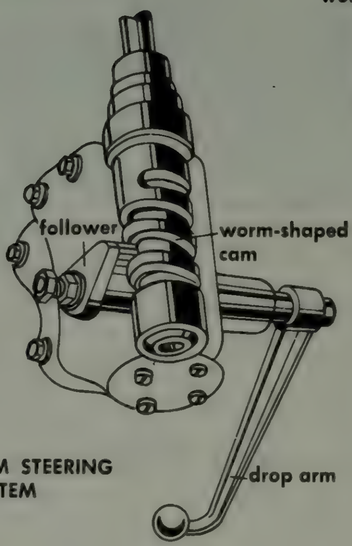


Fig. 5 CAM STEERING SYSTEM

In modern motor vehicles fitted with shoe-type brakes the latter are almost invariably of the internal-expanding kind, hydraulically operated. Hydraulic actuation is based on a natural law, namely, that a pressure exerted upon a liquid is uniformly transmitted in all directions (Fig. 1): the load on the piston at the extreme left is 100 lb.; as a result of the transmission of the pressure through the liquid, which is an incompressible medium, the force developed at each of the other pistons (which are assumed to have the same area as the first) will also be 100 lb. However, the travel of the eight right-hand pistons will be only one-eighth of the travel of the left-hand piston. Fig. 2 shows the diagram of an internal-expanding brake, together with its hydraulic equipment; it comprises a main cylinder with a reserve fluid tank, the wheel cylinders, and the connecting pipelines. Application of the brakes involves the following operations: when the brake pedal is depressed, a piston in the main cylinder is moved (this piston can be regarded as corresponding to the left-hand piston in Fig. 1) and produces a pressure through the brake system. This pressure forces the two small pistons in the wheel cylinders apart. As a result, the shoes are thrust against the brake drum (Figs. 3a and 3b). The kinetic energy of the vehicle is transformed into heat energy by the friction which occurs at the brake linings. This causes heating of the brake drum. In order quickly to get rid of as much heat as possible, the brake drums must have a large external surface area and be mounted in a place where the air has proper access to them. If a brake drum becomes too hot, it will expand excessively, and then the brake shoes will not press so tightly against the inside of the drum. In addition, the braking efficiency of the brake lining diminishes at elevated temperature, as heat reduces the frictional force developed. Thus, poor heat dissipation greatly reduces the efficiency of the brake.

The central component of the brake system is the master cylinder (Fig. 4). When the brake pedal is depressed, the piston travels to the right and produces a pressure in the chamber behind it. This pressure is transmitted by the hydraulic fluid through the pipes to the wheel cylinders. To equalise any differences of pressure in the system (which differences may, for example, be caused by expansion of the brake liquid in the pipelines), a check valve is installed between the pressure chamber and the pipeline. In addition, the master cylinder ensures uniform filling: when the piston is at rest, hydraulic fluid flows from the reserve tank through the compensating port into the pressure chamber. To prevent air being drawn in when the piston returns to the home position, the space behind the back of the piston is kept filled with fluid through the auxiliary port. At one end of the master cylinder is the stop-light switch in which the hydraulic pressure moves a small piston which actuates the electrical contact that switches on the light.

There are various patterns of internal-expanding shoe-type brakes. Figs. 5a and 5b illustrate two patterns of the Simplex brake embodying mechanical arrangements whereby better braking power is obtained by so-called self-servo action.

(Continued)

Fig. 1 PRINCIPLE OF THE HYDRAULIC BRAKE (*Pascal's law*)

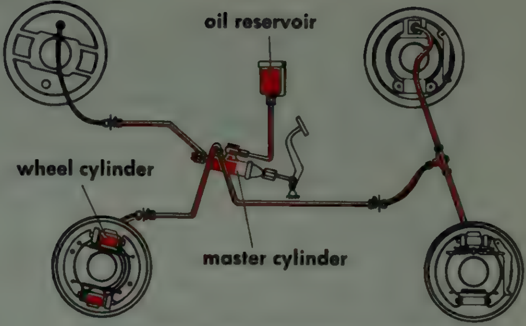


Fig. 2 HYDRAULIC BRAKE SYSTEM (*schematic*)

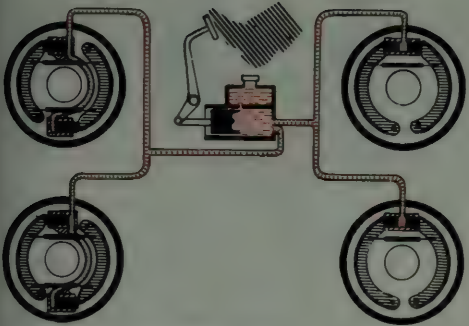


Fig. 3a NEUTRAL POSITION

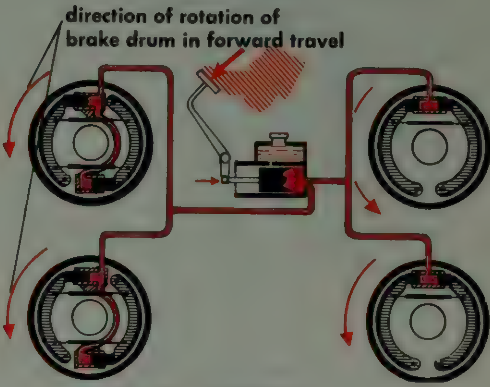
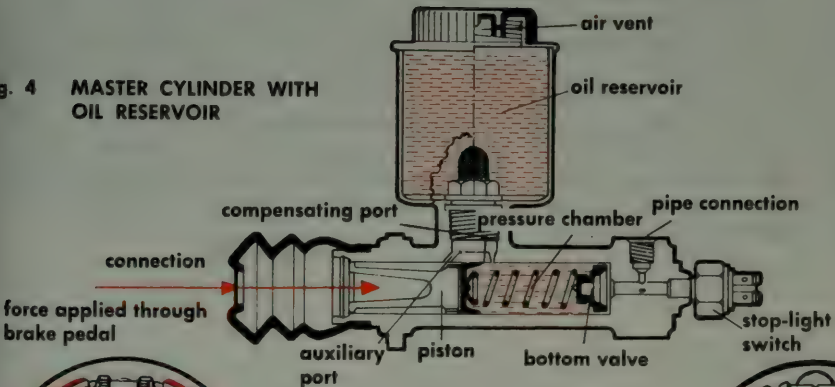


Fig. 3b BRAKING

Fig. 4 MASTER CYLINDER WITH OIL RESERVOIR



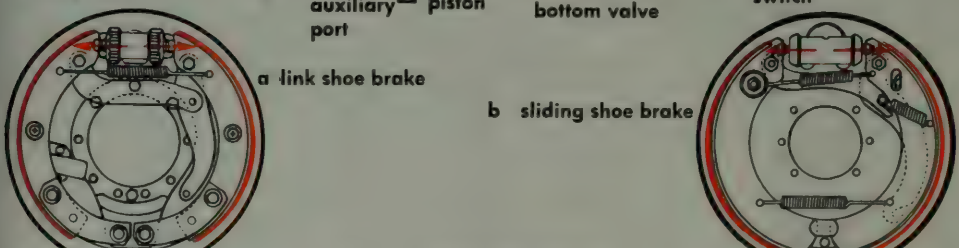


Fig. 5 SHOE-TYPE BRAKES

Another example of the internal-expanding brake is the Duplex brake (Fig. 6): in this arrangement each shoe has its own wheel cylinder which acts in one direction only, pressing one shoe against the drum, while the other shoe serves as a support for thrusting against. In this system the shoes are so mounted that, when the vehicle is travelling in the forward direction, the drum tends to carry each shoe around with it in the direction away from the pivot point of the shoe. This makes for efficient braking by self-servo action. On the other hand, the braking action is rather poor when the vehicle is reversing.

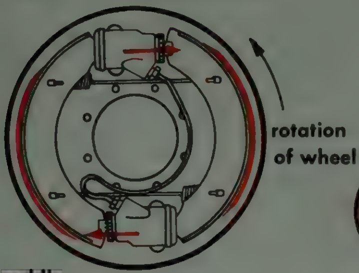
Besides the single-circuit brake systems described here, dual-circuit hydraulic systems comprising a tandem master cylinder which contains two pistons that are actuated by the brake pedal (Fig. 7). The front piston transmits its force to the rear one, so that hydraulic pressure is developed in two circuits. The advantage of the two-circuit braking system is that if an oil pipe in one circuit fractures, the second circuit will still continue to function.

Disc brakes

Partial disc brakes: In this arrangement the brake disc is gripped pincer-wise between two circular or kidney-shaped brake pads fitted with friction linings. The large area of the disc assures excellent dissipation of the heat generated by braking (Figs. 8a and 8b). Two-circuit hydraulic systems, when applied to disc brakes of this kind, do not separately serve the front and the rear wheels respectively; instead, each circuit serves all four wheels. For this purpose the brake disc of each of the wheels is provided with four hydraulic brake cylinders (Fig. 9).

Full disc brakes (Fig. 10): In this form of construction, fixed pads with friction linings act from inside upon both sides of a rotating casing, which has to dissipate the heat in much the same way as a shoe-type brake. In this case the self-servo or self-energising action (Fig. 10b) is obtained by means of steel balls which mount the inclined faces of sockets when the two brake discs (one fixed and one moving) rotate in relation to each other. This forces the discs apart, causing them to be pressed even more firmly against the casing. The degree of servo action developed is dependent on the inclination of the faces of the sockets. If they are flatly inclined, a considerable servo action results, but this may involve a risk of jamming and locking of the brake, as the balls may be wedged so tightly that they cannot by themselves roll back to the bottom of the sockets. Besides, the two discs must then undergo a considerable displacement in relation to each other before they are forced so far apart that they can develop a powerful braking action. Wear of the friction linings necessitates an even greater amount of displacement of the discs in relation to each other. For this reason disc brakes having a high degree of self-servo action are particularly sensitive to lining wear.

Fig. 6 DUPLEX BRAKE



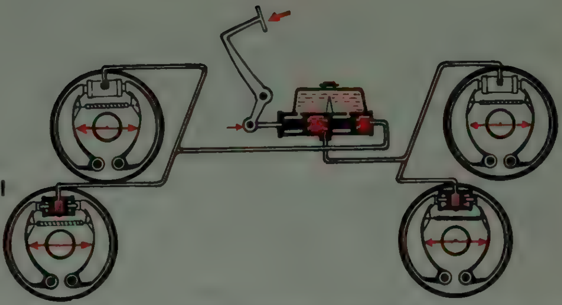


Fig. 7 DUAL-CIRCUIT HYDRAULIC SYSTEM WITH TANDEM MASTER CYLINDER

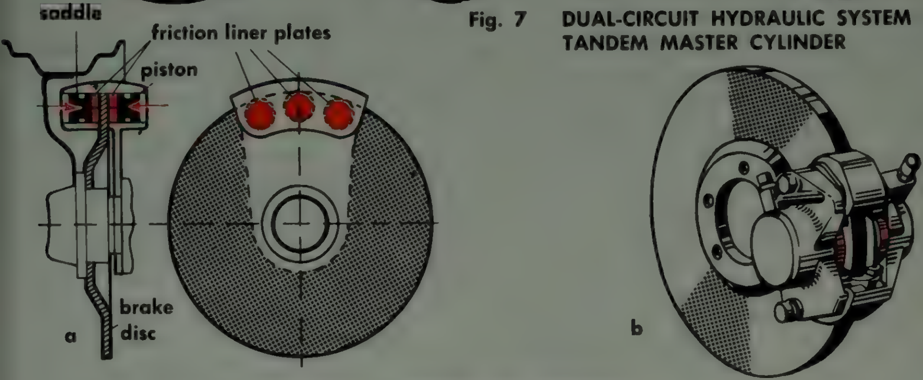


Fig. 8 PARTIAL DISC BRAKE

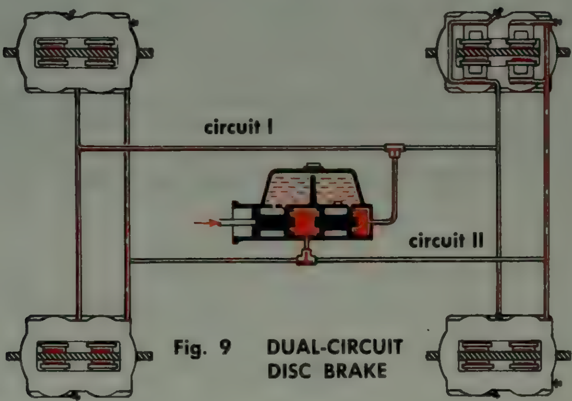


Fig. 9 DUAL-CIRCUIT DISC BRAKE

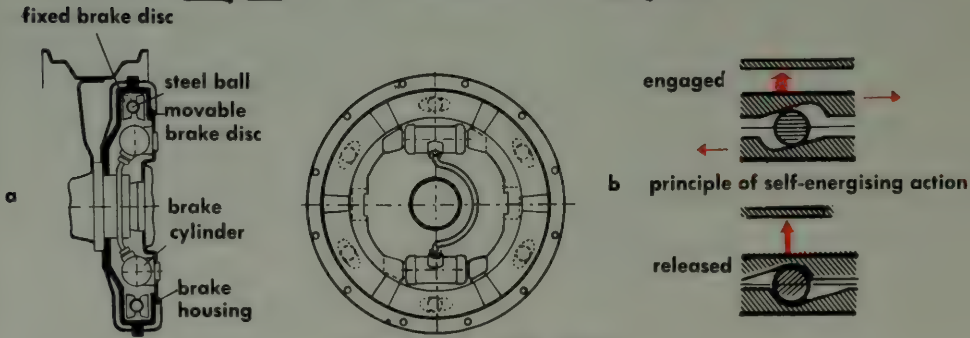


Fig. 10 FULL DISC BRAKE

A car headlamp consists of a housing (usually recessed into the body of the vehicle), the reflector, the light-diffusing glass, and the bulbs with their sockets.

The purpose of the reflector is to concentrate the light that the bulb emits in all directions and directing it ahead of the vehicle. A reflector is a paraboloid mirror which is so shaped that all the rays of light emitted from a light source at its focus are reflected in a direction parallel to the axis of the reflector ((Fig. 1;. p. 160,gvol. I) If the light source is not located at the focus (Figs. 2 and 3), the light output of the headlamp in the desired direction will be greatly reduced by scatter of the rays. The reflectors are made of steel, the reflecting surface being "silvered", usually by a thin coating of aluminium applied as vapour in vacuum. They reflect about 89% of the light that strikes them. The headlamp glass functions as a light diffuser, i.e., it must so distribute the light that the zone on each side of the vehicle and directly in front of it must also be illuminated. This effect is achieved by providing the glass with grooves or ribs which function as prisms (Fig. 4) and deflect the light. This diffusion does, of course, reduce the brightness of the beam in the main direction.

The principal requirements that the electric bulbs have to fulfil are that the filament must be accurately located at the focus of the reflector and that the bulb should be as small as possible and yet be very bright. The first requirement is fulfilled by accurate location of the fixing tabs on the base of the bulb in relation to the filament. Thus, if the bulb is correctly fitted, the filament will be exactly at the focus of the reflector. A high light efficiency is obtained with coiled or "coiled coil" tungsten filaments enclosed in bulbs filled with an inert gas.

At the present time most headlight bulbs are of the twin-filament type, with a built-in screen (Figs. 5a and 5b). The filament for high-beam ("far beam") is located at the focus of the reflector. The filament for low beam ("dimmed light") is placed a few millimetres in front of the focus and a little higher (Fig. 5a), so that the light emitted by this filament is reflected as a downward-directed spreading beam (Fig. 6a). Rays from this filament striking the lower part of the reflector would be reflected upwards, but these rays are intercepted by the screen installed (inside the bulb) under the filament.

Most cars in the United States are equipped with sealed-beam headlights. This type of light contains the filament, reflector, and lense, all assembled into one unit. The entire sealed-beam fits into the car's headlight housing and, if replacement becomes necessary, the entire unit is removed and a new one is inserted.

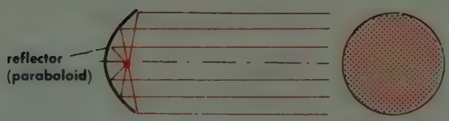


Fig. 1 LIGHT SOURCE AT FOCUS:
rays are reflected as a parallel beam

Fig. 2 LIGHT SOURCE ABOVE OR BELOW
FOCUS: rays are scattered

Fig. 3 LIGHT SOURCE BEFORE OR BEHIND FOCUS:
rays are scattered

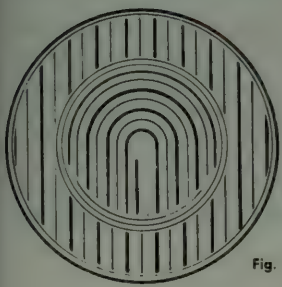


Fig. 4a LIGHT-DIFFUSING GLASS

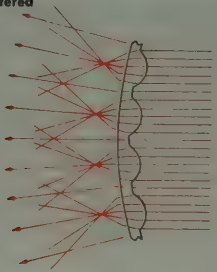


Fig. 4b LIGHT DISTRIBUTION
BY DIFFUSING GLASS

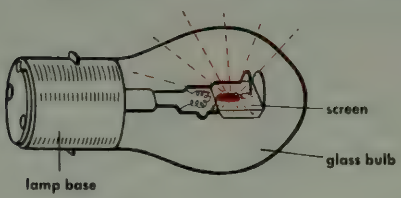


Fig. 5a TWIN-FILAMENT LAMP
(dimmed light)

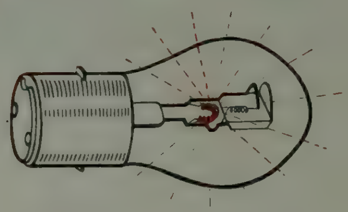


Fig. 5b TWIN-FILAMENT LAMP
(far beam)

Fig. 6a DIRECTION BEAM WITH DIMMED LIGHT

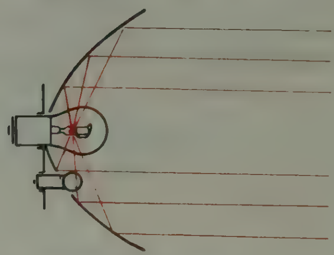


Fig. 6b DIRECTION OF FAR BEAM

A motor horn, like any other source of sound, must set up vibrations in the air; the variations in air pressure associated with these vibrations are perceived by the ear as sound. The pitch of the sound is determined by the frequency, i.e., the number of vibrations per second. Depending on the acoustic pressure (measured in microbars, a unit of pressure corresponding to a water volume of 1/100 mm) and the frequency of the sound, the human ear perceives a certain loudness (Fig. 1). Loudness is measured in phons. For a given acoustic pressure, sounds differing in pitch are not perceived as having the same loudness. The range of audibility is bounded by the threshold of hearing at low values of the acoustic pressure, and by the threshold of pain at high values of this pressure. At frequencies of 2000–5000 cycles/secs., audibility extends down to very low acoustic pressures, i.e., the human ear is particularly sensitive in this frequency range. It so happens that the acoustic pressure of the sound emitted by motor vehicles is relatively low in this same frequency range of 2000–5000 cycles/sec. (see Fig. 2). It is therefore advantageous to ensure that the frequencies of the sounds made by a motor horn are within this range: the horn will thus better be able to make itself heard above the general noise of the traffic. Most of the horns now used as warning devices on motor vehicles are either of the impact type (Fig. 3) or klaxons (Fig. 4). Both kinds have an electromagnet whose winding is energised when the driver presses the horn button. An iron plate, called the armature, is attracted by the magnet and, in doing so, opens a contact and thus cuts off the current to the electromagnet, which now releases the armature, allowing the latter to spring back to its initial position. The whole cycle of events now repeats itself. The oscillations of the armature are transmitted to a diaphragm which in turn sets up vibrations in the air. The number of vibrations per second is called the fundamental frequency of the horn.

In the impact-type horn (Fig. 3) the entire diaphragm assembly (comprising armature plate, oscillating beam and disphragm) strikes the core of the electromagnet. Harmonic oscillations in a narrow frequency range are additionally produced. As a result of this, the sound spectrum of an impact-type horn mainly shows two narrow frequency ranges, so that it can penetrate effectively through noise. In the case of the klaxon, with its rather softer note, the diaphragm assembly does not hit the core, but can oscillate freely. The sound spectrum of the klaxon has many overtones spread out over a wide frequency range; the pitch of the fundamental tone of the horn is determined mainly by the length of the horn.

Fig. 1 CURVES OF EQUAL LOUDNESS SHOW THAT THE SENSITIVITY OF THE HUMAN EAR IS GREATEST BETWEEN 2000 AND 5000 CYCLES PER SECOND

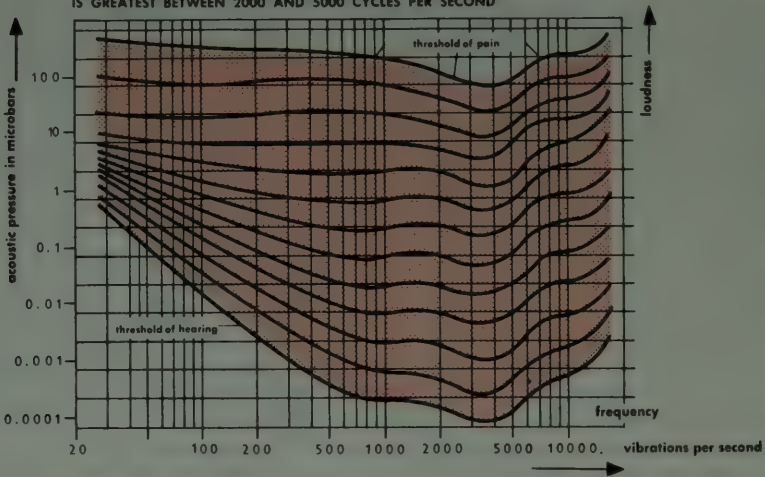


Fig. 2 NOISE OF MOTOR TRAFFIC HAS LOW ACOUSTIC PRESSURE BETWEEN 2000 AND 5000 CYCLES PER SECOND

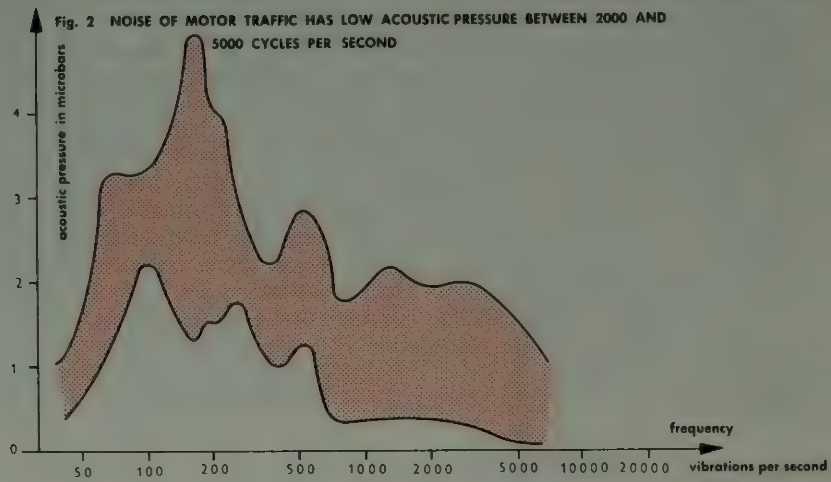


Fig. 3 IMPACT-TYPE HORN

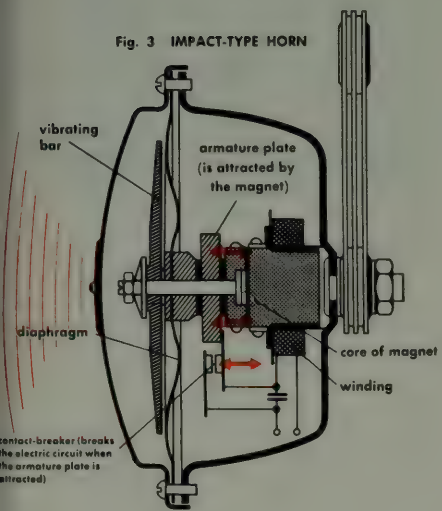
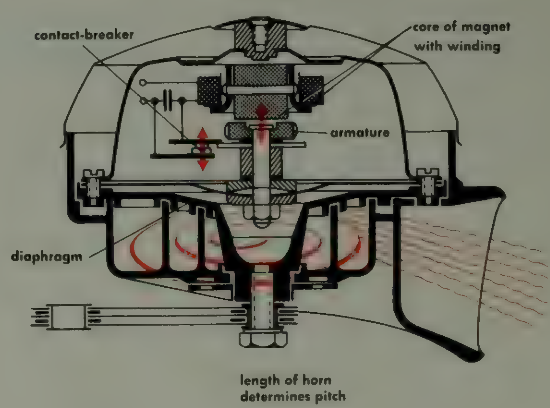


Fig. 4 CLAXON



The flashing direction indicator (or blinking trafficator) operates on the electro-thermal principal and uses bimetallic strips to open and close electrical contacts. As its name implies, a bimetallic strip consists of two strips of different metals, with different coefficients of thermal expansion, bonded together. When a bimetallic strip is heated, it will curve towards the side where the metal with the lower coefficient of expansion is affixed. Thus, if one end of the strip is gripped, the other end will thus move a certain distance away from its home position when cold, the magnitude of this distance depending on the temperature to which the strip is exposed.

The flashing direction indicator comprises two such bimetallic strips (Fig. 2a), both of which move in the same direction on being heated. Each strip is provided with a silver contact at its end. In the neutral position these contacts touch each other. When the ignition of the motor vehicle is switched on, current from the battery flows through these contacts and through the heating coil of one of the bimetallic strips. The heat produced by this coil warms the strip, causing it to bend and thus open the contact. Now the current, instead of flowing through this contact, is passed through the heating coil of the other bimetallic strip, with the result that this strip, too, is deflected, in the same direction as the other strip (Fig. 2b). The latter now receives less current, consequently cools a little, and moves back towards its initial position; the two contacts touch each other again; and the cycle is repeated. This continues as long as the ignition is switched on, even while the direction indicator switch is not actuated by the driver of the vehicle. When he wishes to turn a corner and actuates the switch, then current will flow directly to the direction indicator when the contacts are closed (Fig. 2a). The indicator lamp then lights up. However, the process continues: the contacts separate, and the lamp goes out. The pilot light inside the vehicle always lights up just when the indicator light is out because the pilot light gets its current only when the contacts of the bimetallic strips are open (Fig. 2b). When these contacts close again, the indicator lamp lights up, while the pilot light remains dark, as it is then short-circuited by the contacts. The frequency of blinking is independent of the outside temperature. Besides, when the engine is running, the flashing indicator is already switched on and therefore comes into immediate action when the driver actuates the indicator switch.

Another widely used type of flashing indicator is based on the hot-wire principle. The thermal elongation of a thin resistance wire under the influence of a current passing through it is used for controlling the blinker contacts.

Fig. 1

PARTS OF A FLASHING DIRECTION INDICATOR

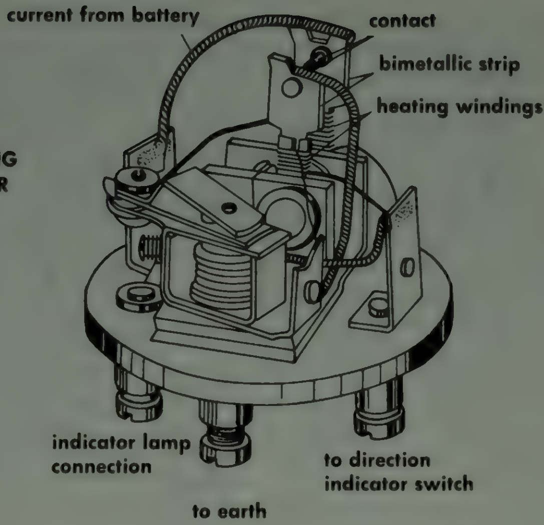


Fig. 2 a

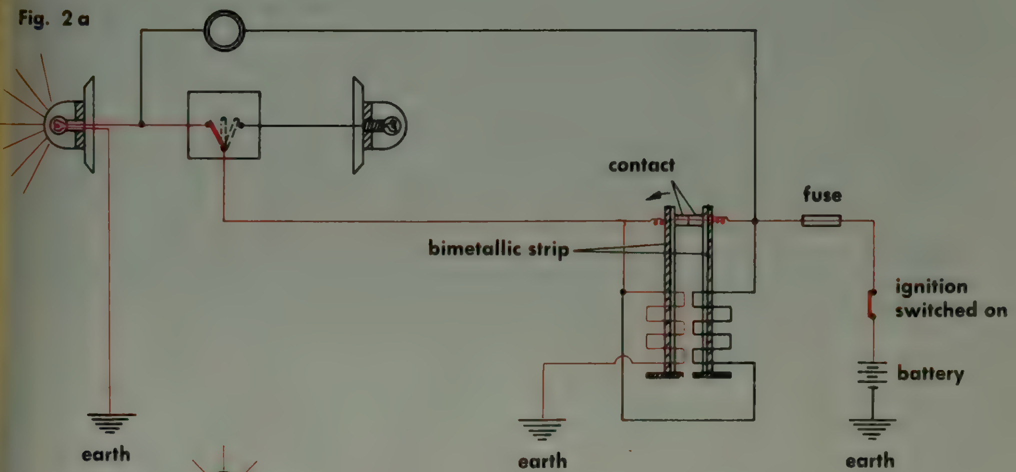
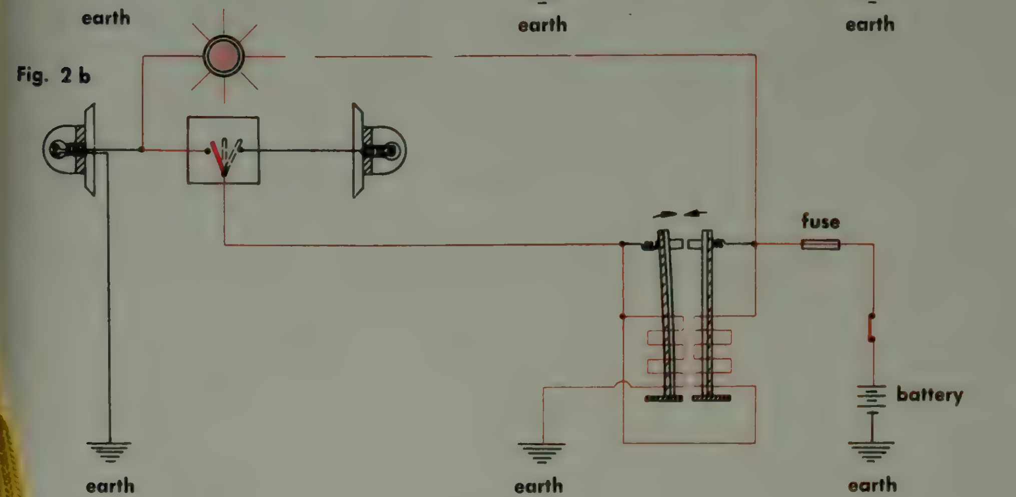


Fig. 2 b



SPEEDOMETER

A speedometer is an instrument which measures the speed at which the car is travelling and usually also embodies a mileage recording mechanism. The central feature of the device is a permanent magnet. Each magnet is surrounded by a magnetic field (Fig. 2), which can be conceived as consisting of lines of force. These can be "made visible" by strewing iron filings on a sheet of cardboard under which the magnet is held; the filings will arrange themselves along the lines of force and thus reveal the pattern of these lines. When a magnet is rotated, its field will rotate with it. The magnet in the speedometer begins to rotate as soon as the vehicle is set in motion; it is driven through a small gear unit by the speedometer shaft, which is connected to the propeller shaft or the front axle. The higher the speed of the vehicle is, the higher is the speed of rotation of the magnet. The magnet rotates concentrically in an aluminium ring, in which the rotating magnetic field induces eddy currents which in turn produce a magnetic field of their own. The interaction of this magnetic field with that of the rotating permanent magnet exercises a torque (twisting moment) on the aluminium ring. This torque tries to rotate the ring along with the magnet. The faster the magnet rotates, the higher is the torque. The ring is not free to rotate, however: it can merely swing a certain distance—depending on the magnitude of the torque—and is then restrained by the counteracting force of a spiral spring. Attached to the ring is a pointer which indicates the speed of the vehicle on a suitably graduated scale. In many types of speedometer the ring has an extension in the form of a drum to which the scale is attached. When the speed increases, this drum rotates, and the speed indication is shown in a narrow gap in a panel mounted in front of the scale. The speedometer generally also comprises a mileage recorder. This device is driven through a small worm gear, which is mounted on the speedometer shaft, by the same shaft that drives the rotating magnet. The motion of the speedometer shaft, greatly reduced by another small gear unit, is transmitted to the mileage recording mechanism. The latter therefore merely counts the number of revolutions. However, since a certain number of revolutions, of the front wheel or of the propeller shaft correspond to a certain distance travelled by the vehicle, the revolution counting mechanism can be made to give a direct reading of the mileage covered. When the disc which counts the units (i.e., miles) has performed one whole revolution, the counting disc for the tens is rotated one place (by means of a small driving catch). When the "tens" disc has performed one revolution, it similarly rotates the "hundreds" disc one place. This operation is repeated up to the "ten thousands" disc.

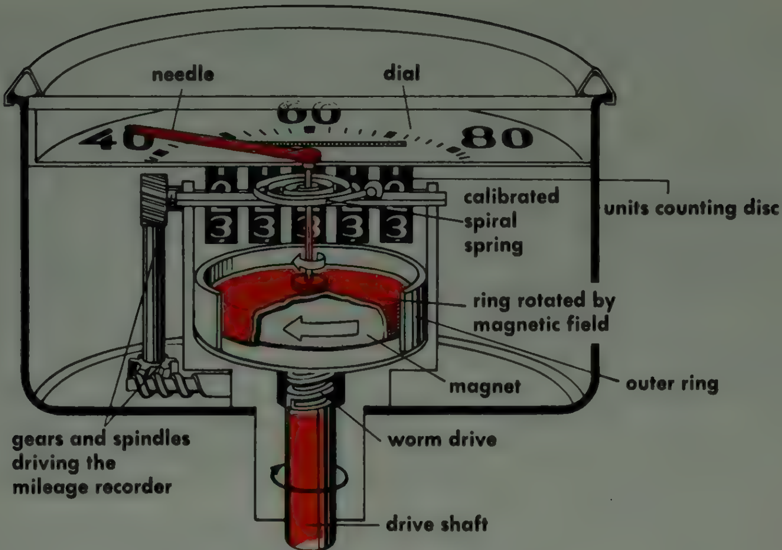


Fig. 1 SPEEDOMETER
(*schematic*)

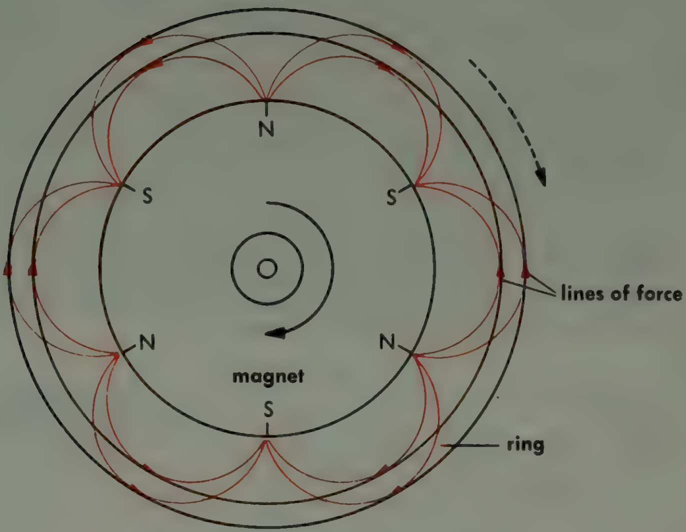


Fig. 2 PRINCIPLE OF SPEEDOMETER DRIVE

A steam locomotive is powered by a reciprocating steam engine, i.e., a power unit operated by the expansion of steam which is admitted into a cylinder and moves a piston to and fro. The piston motion is transmitted to the driving wheels of the locomotive.

The drive mechanism and control gear of a steam locomotive are illustrated in Fig. 1. The steam, generated in a smoke tube boiler or flue boiler (p. 50, vol. I and Fig. 3), is delivered through a pipe to the slide-valve chest. From here the slide valve admits the steam to the cylinder alternately through two openings so that it enters first on one side and then on the other side of the piston and thus pushes the latter to and fro. In Fig. 1 the steam is entering the cylinder on the left-hand side of the piston, and is pushing the piston to the right. At the same time, the exhaust steam on the right-hand side is being pushed out of the cylinder by the piston and is discharged through a duct from which it excapes through the chimney into the open air. The piston rod connects the piston to the cross-head, which moves backwards and forwards in a guide. Attached by means of an articulated connection to the cross-head is the driving rod whose other end is connected to the driving pin on the driving wheel. In this way the reciprocating (to-and-fro) motion of the piston is transformed into the rotary motion of the wheel. There are usually two or more sets of driving wheels, these being coupled together by coupling rods.

The locomotive in Fig. 1 is travelling backwards; its driving wheels are revolving anti-clockwise. To reverse the motion, the slide valve would, for the piston position illustrated, have to be so altered so as to admit steam to the right instead of to the left of the piston. This adjustment of the slide valve can be done from the driver's cab by means of a handwheel which works a system of control rods and thus pulls the slide valve to the left. This adjustment system is shown diagrammatically in Fig. 2. The normal to-and-fro movement of the slide valve to admit the steam alternately on the two sides of the piston is performed by the slide valve gear. Attached to the same pin as that actuated by the driving rod is a so-called fly crank which, through a rod, moves the link to and fro. When the valve rod is above the pivot, the slide valve will move to the left when the link rod moves to the right; when the valve rod is below the pivot, the opposite occurs.

The locomotive develops its tractive force because of the "adhesion" (due to friction) between the wheels and the rails. Because of the weight of the locomotive, whereby the wheels are pressed tightly against the rails, a high frictional force and therefore a high tractive force can be developed. However, if the force applied to the wheel circumference by the steam engine exceeds the frictional force that can be developed, slip of the wheels on the rails will occur.

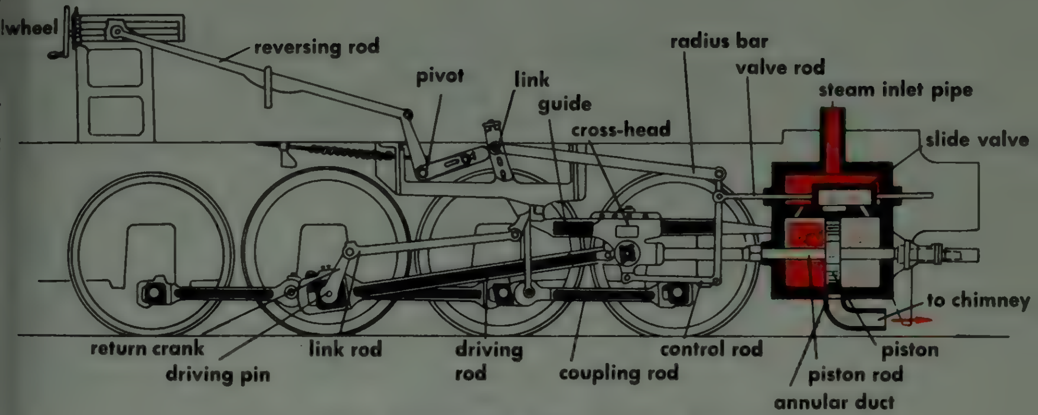


Fig. 1 DRIVE MECHANISM AND CONTROL GEAR OF A STEAM LOCOMOTIVE

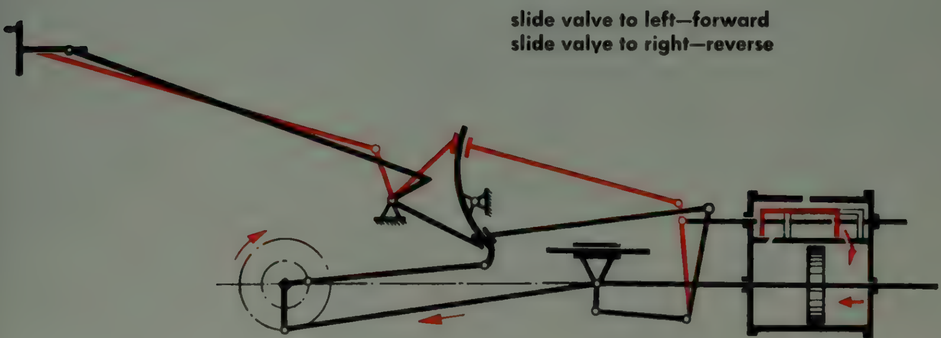


Fig. 2 DIAGRAM SHOWING SLIDE VALVE ADJUSTMENT
(*Heusinger system*)

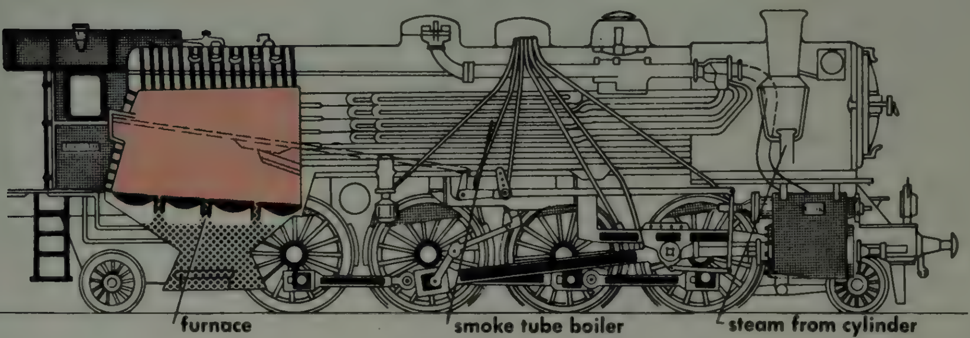
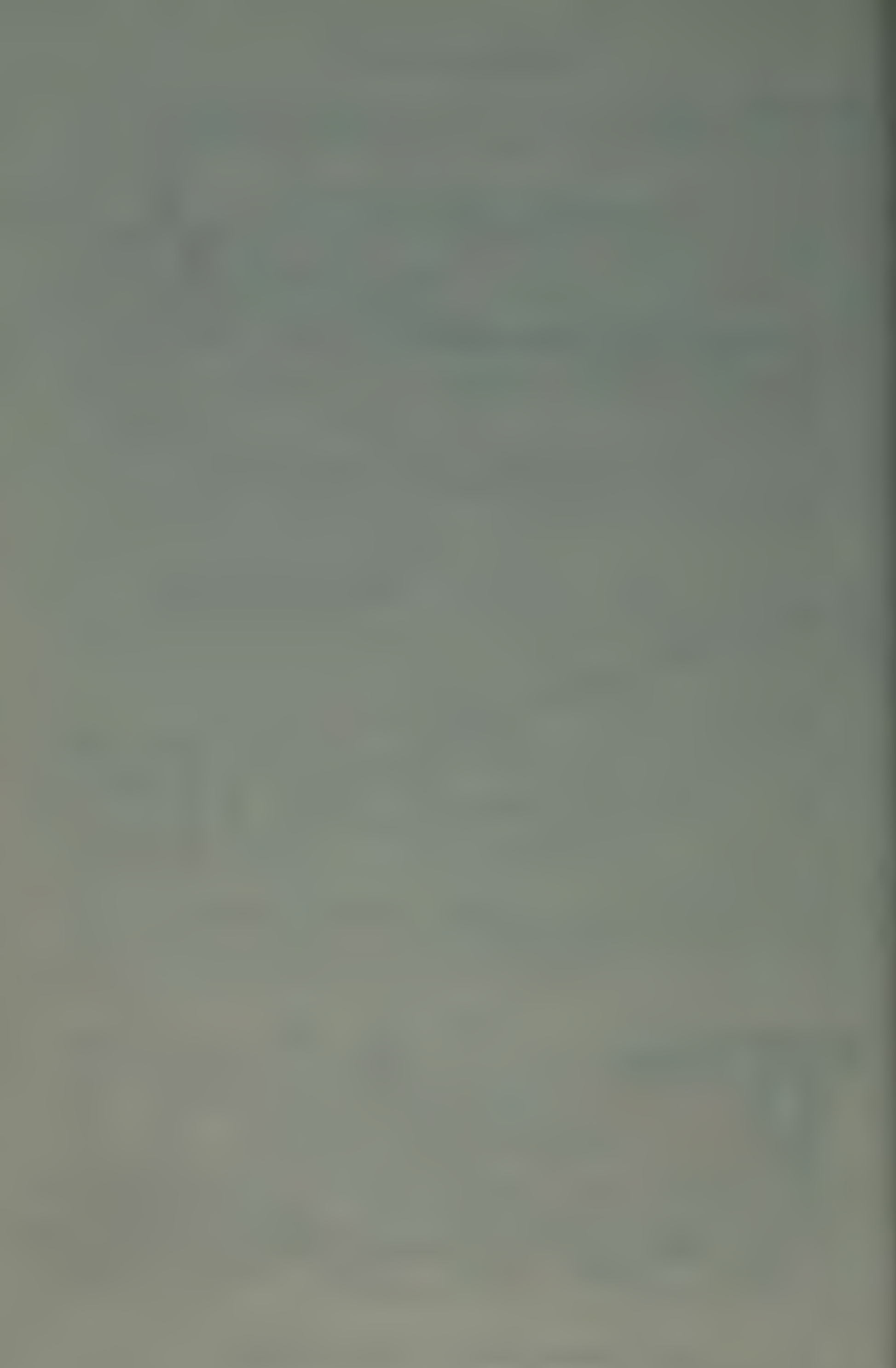


Fig. 3 SECTION THROUGH A STEAM LOCOMOTIVE



Steam locomotive, top speed: 63 m.p.h
Photo CFF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As a traction unit for railway operation the steam locomotive (see page 248) has, in many countries, been superseded by the electric locomotive. One of the advantages of the latter is that the power to drive the locomotive does not have to be generated in the locomotive itself but can be supplied to it through overhead wires or through conductor rails. A further advantage is that the electric motor develops its highest torque (turning moment applied to the wheels) at starting. This enables the locomotive, and the train it pulls, to move off more swiftly after stopping at a station. Once the locomotive is in motion, the torque required to keep it going is less and the amount of electric power consumed is accordingly reduced.

The wheels or the axles can be individually driven by a separate electric motor to each axle (Fig. 1). Alternatively, one large motor may drive a number of wheels through a system of connecting rods (Fig. 2). With individual drive, relatively small high-speed motors are employed whose speed is reduced to the drive speed of the wheels by means of gears. Large motors can more suitably be designed to run at lower speeds and directly drive an intermediate driving shaft (or jackshaft). Fig. 3 represents an electric motor with individual drive. The current is collected from the overhead contact wire by means of pantograph current collectors. On many railways single-phase alternating current at 15,000 volts with a frequency of $16\frac{2}{3}$ cycles per second is used. This current is transformed down to the working voltage of the motors (usually 300–700 volts). The speed of the locomotive is controlled by connecting the motors to different tappings (output connections) of the transformer (cf. page 110, vol. I). This is preferable to controlling the speed by means of variable resistances connected in series with the motors, as the latter method would entail considerable power losses.

A variant of the electric locomotive is the diesel-electric locomotive, which generates its own electric current by means of a generator (p. 78, vol. I). powered by a diesel engine (see page 194). This makes the locomotive independent of an outside power supply. Diesel-electric locomotives are used chiefly on non-electrified or only partially electrified railway lines.

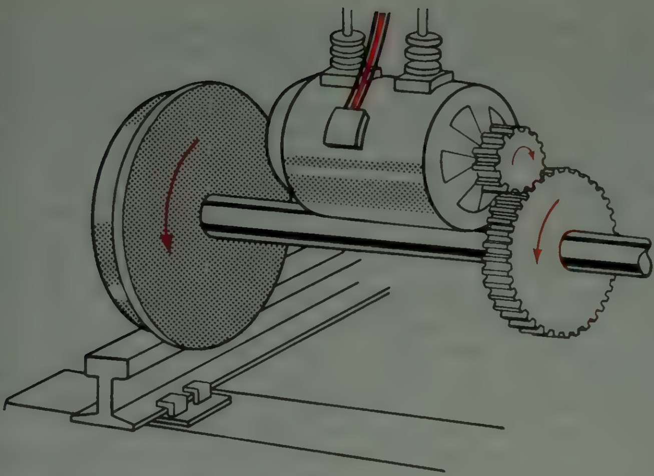


Fig. 1 INDIVIDUAL AXLE DRIVE

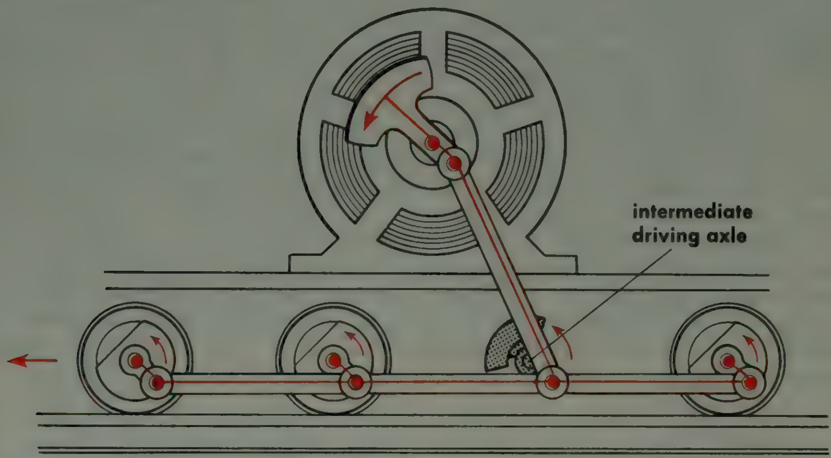


Fig. 2 MULTIPLE-AXLE DRIVE BY ONE MOTOR THROUGH CONNECTING ROD

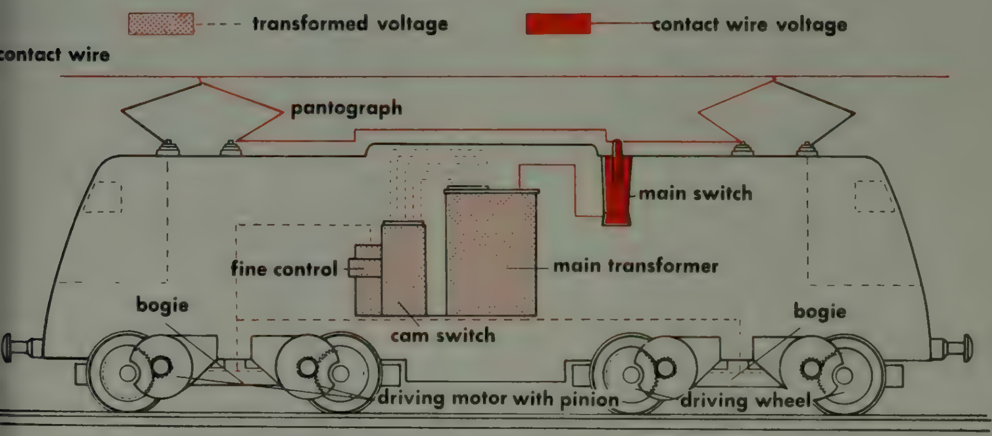


Fig. 3 SCHEMATIC SECTION THROUGH AN ELECTRIC LOCOMOTIVE

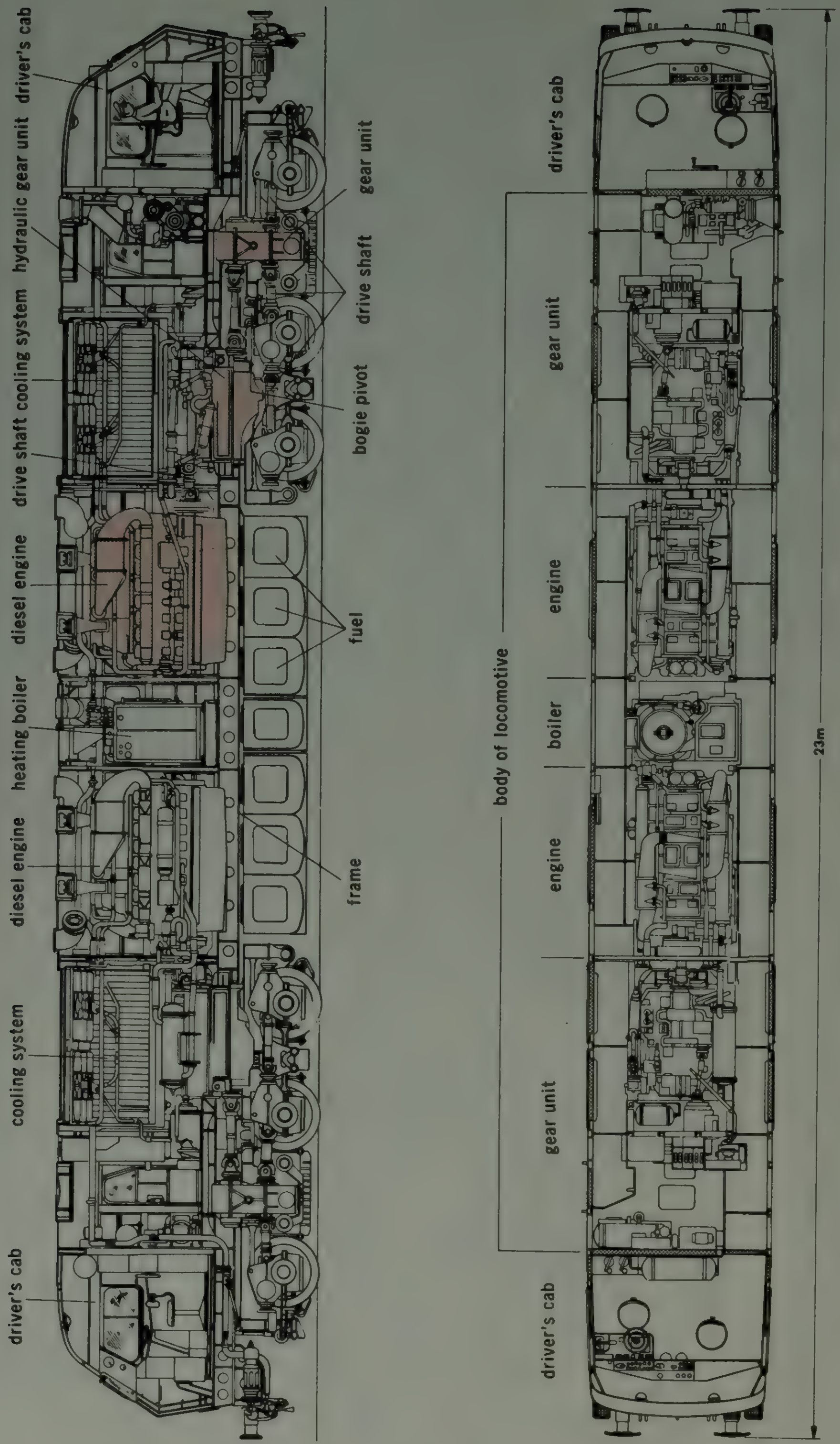
Diesel locomotives are relatively cheap to run, can be got ready for operation and started up more quickly and conveniently than steam locomotives, and can be speedily and easily refuelled. Also, their mechanical efficiency is superior to that of steam locomotives, while they dispense with the expensive overhead contact wires and power distribution installations required by electric traction. However, they lack the robustness of steam locomotives and the operational flexibility and smooth running properties of electric locomotives.

The accompanying illustration shows a typical modern diesel locomotive, the type V 320 as used by the West-German Federal Railways. This locomotive has two independent sets of drive machinery, each associated with one bogie (or truck). The power is supplied by two 16-inch cylinder four-stroke diesel engines developing 2000 h.p. each. These are supercharged engines, i.e., the air needed for combustion of the fuel is fed to the cylinder by powerful fans (radial-flow fans) driven by small gas turbines (p. 56, vol. I) which utilise the exhaust gases from the engine. Supercharging greatly increases the power of an engine in comparison with a non-supercharged engine which has to rely on mere suction for its air intake.

The engine power is transmitted through rubber-cum-metal couplings and shafts to two hydraulic gear units and thence through another shaft to a special mechanical gear unit which can be hand-controlled from low to high speed range. From this speed-control gear unit all the axles of the bogie are driven by means of a number of drive shafts. The engine itself as well as the lubricating oil and the gear oil are cooled by water which is then circulated through a fan-operated cooling unit where it gives off the heat it has collected.

The underframe and superstructure of the V 320 are of lightweight welded steel construction; the superstructure shell is welded to, and co-operates structurally with, the underframe. The locomotive is equipped with disc brakes—a pair of brakes to each wheel—and additionally an electro-magnetic rail brake which ensures short stopping distances even at high speeds.

The electrical equipment of the locomotive comprises two electric generators which supply current to the lead storage batteries, the motors for driving various items of auxiliary equipment, the lighting system, and the electric control and monitoring equipment. In addition, the V 320 diesel locomotive is provided with a boiler installation for supplying superheated steam for heating the train and for preheating the engine.


driver's cab
hydraulic gear unit
cooling system
drive shaft
diesel engine
heating boiler
diesel engine
cooling system
driver's cab
gear unit
drive shaft
bogie pivot
fuel
frame
driver's cab
gear unit
engine
boiler
engine
gear unit
driver's cab
body of locomotive
23m

Diesel-electric locomotive, top speed: 47 m.p.h
Photo CFF

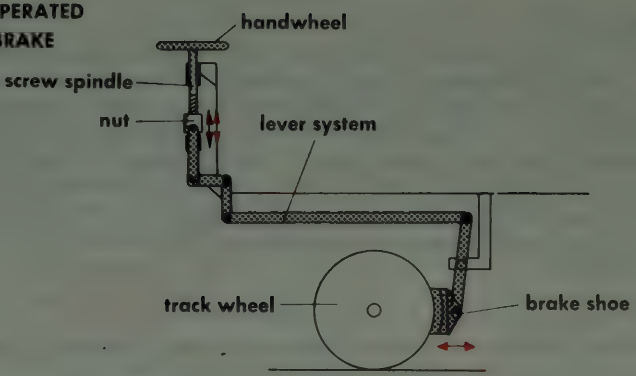
The simplest method of braking a railway vehicle is by means of the hand-operated block brake (Fig. 1). The braking force in this case is applied through a system of levers worked by a handwheel mounted on a screw spindle.

For quickly stopping a train in motion, however, it is necessary to employ more powerful means, in the form of a compressed air brake (pneumatic brake) as illustrated in Figs. 2a and 2b. The system comprises a compressor which supplies compressed air to the main air reservoir and thence, via the driver's brake valve and the main brake pipe, to the auxiliary air reservoirs. The main brake pipe extends through all the carriages of the train, the connections between the carriages being formed by flexible hoses provided with quick-action couplings. When the brake is released (Fig. 2a), there is equal pressure on both sides of the valve piston. Its own weight keeps this piston pressed down on its seat and thus closes the passage through which the compressed air can reach the brake cylinder, which is in communication with the external air. Each axle is equipped with a brake cylinder whose piston actuates the brake.

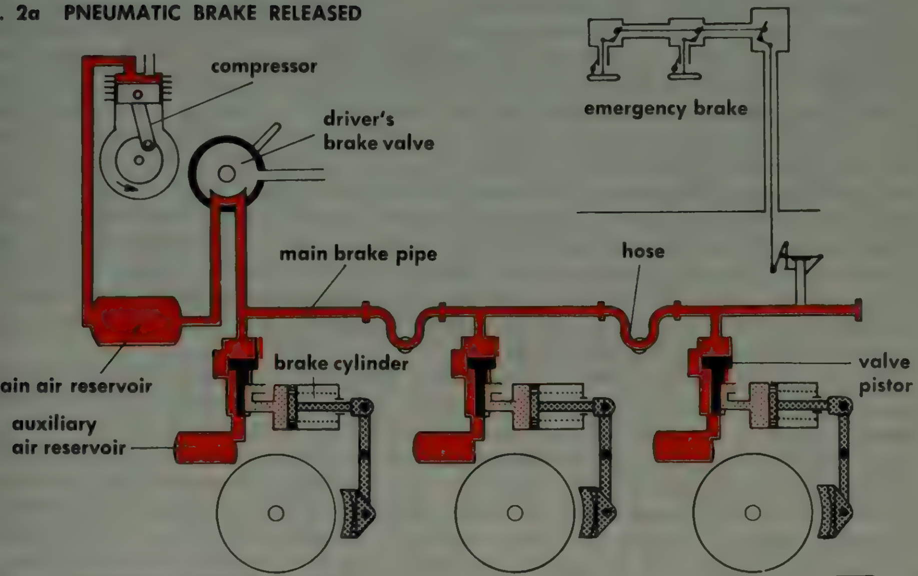
When the driver wishes to apply the brakes, he moves the control lever of the brake valve to the appropriate position (Fig. 2b). As a result, the main brake pipe is put into communication with the external air, and the pressure in this pipe therefore goes down to the atmospheric pressure. Now all the auxiliary reservoirs contain air at a higher pressure than that in the main brake pipe. This causes the brake valve piston to be forced upwards, whereby compressed air, from the auxiliary reservoirs, is admitted to the brake cylinder. The compressed air entering this cylinder pushes the brake piston back against the pressure of a spring (which keeps the piston in the home position when the brake is released) and thus causes the brake blocks to be pressed hard against the wheels. On completion of the braking operation the control lever of the driver's brake valve is returned to its initial position. A slight excess pressure in the main brake pipe pushes the valve piston down again. The main brake pipe is thus reconnected to the auxiliary air reservoirs and the brake cylinders are once again in communication with the atmosphere.

In the event of a connecting hose between two carriages rupturing or bursting, the main brake pipe will lose its air pressure. As a result, the brakes will be applied all along the train, The same effect is produced by actuation of the emergency brake valve which is provided in each carriage and which can be worked from each compartment by means of a control wire. The type of pneumatic brake described here is named the Westinghouse brake, after its inventor. A further development of this system is embodied in the Kunze-Knorr brake, which has a brake cylinder comprising two chambers and two coaxial pistons. This arrangement enables the brake to be released step by step and thus makes for finer brake control.

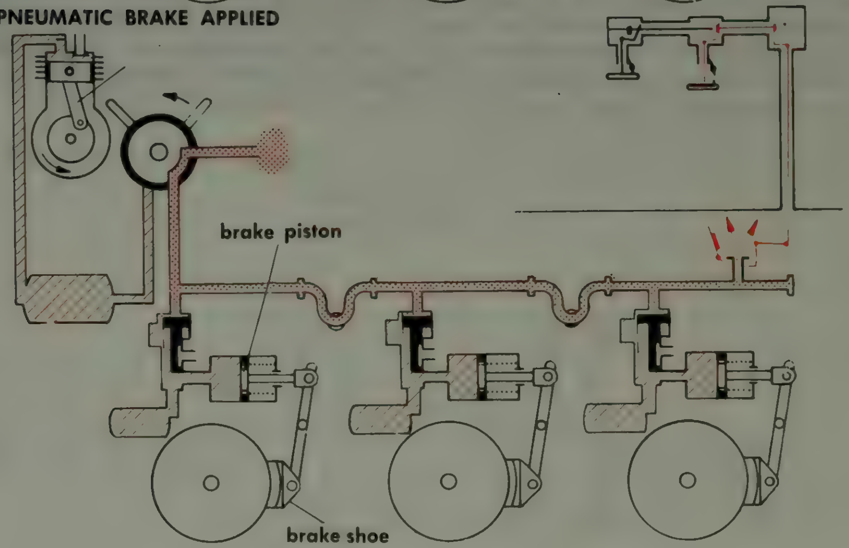
. 1 HAND-OPERATED BLOCK BRAKE



. 2a PNEUMATIC BRAKE RELEASED



. 2b PNEUMATIC BRAKE APPLIED



Automatic train stop:

Attached to the driver's control lever in a railcar or an electric locomotive is a contact which, on being depressed, breaks an electric circuit (Fig. 1). If the driver lets go of the lever, this circuit is closed, causing current to flow through the windings of an electromagnet of a monitoring valve. This valve closes the pipe leading from the main air reservoir to the main brake pipe. At the same time another valve is opened whereby the pressure is released from the main brake pipe, thus causing the brakes to be applied (see page 258), the effect being the same as if the brake valve had been actuated. When the driver's control lever is released, the driving motor is also automatically switched off.

Inductive control:

For use on busy lines various other safety devices are additionally available. One of these is the inductive system (Fig. 2) of train control. Each signal on a section of line equipped with this system has a device whereby the brakes of a passing train can be actuated. The locomotive is provided with a "transmitter" which comprises an iron-core coil and a condenser. This oscillating circuit (p. 86, vol. I) is tuned to 2000 cycles/sec. and is fed with current from a high-frequency generator to which, in addition, the coil of an electromagnet for a relay (p. 112, vol. I) is connected. The high-frequency generator is driven by an electric motor or a small steam turbine which also drives a 24-volt direct-current dynamo. Connected to this dynamo via the high-frequency relay is a second relay and two pilot lamps. When the system is in operation, the high-frequency generator supplies current to the oscillating circuit of the locomotive and to the high-frequency relay, so that the armature of the latter is attracted by the electromagnet and allows current to flow through the 24-volt direct-current relay. When thus energised, this last-mentioned relay interrupts a flow of current to the actuating electromagnet of a valve (corresponding in its function to the monitoring valve in Fig. 1).

The signal beside the railway track is also equipped with an oscillating circuit tuned to 2000 cycles/sec. When the signal is at "safe", its oscillating circuit is short-circuited and has no effect on the passing locomotive. However, if the signal is at "stop" and the train nevertheless fails to stop, the oscillating circuit of the signal will resonate (p. 232, vol. I) in tune with that of the passing locomotive and thereby withdraw so much energy from the latter circuit that the armature of the high-frequency relay is no longer sufficiently strongly attracted by the electromagnet and is therefore released. This in turn cuts off the supply of current to the 24-volt relay, which then likewise releases its armature. This closes the electric circuit to the electromagnet that controls the monitoring valve. When this happens, the monitoring valve causes the brakes to be applied and the driving motor of the locomotive to be switched. At the same time, the pilot lamp "a" goes out and the lamp "b" lights up.

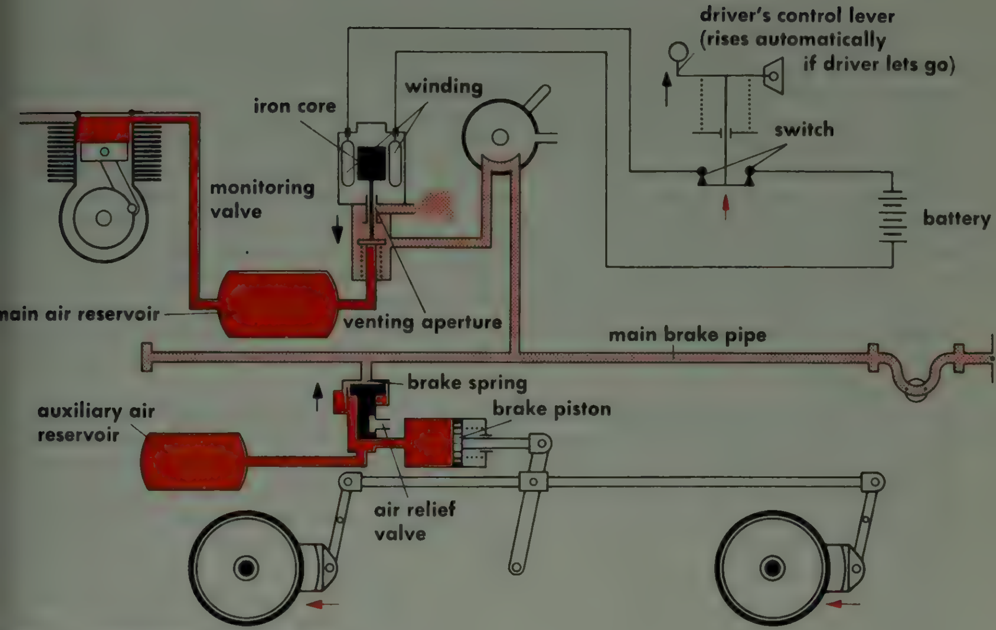


Fig. 1 AUTOMATIC TRAIN STOP

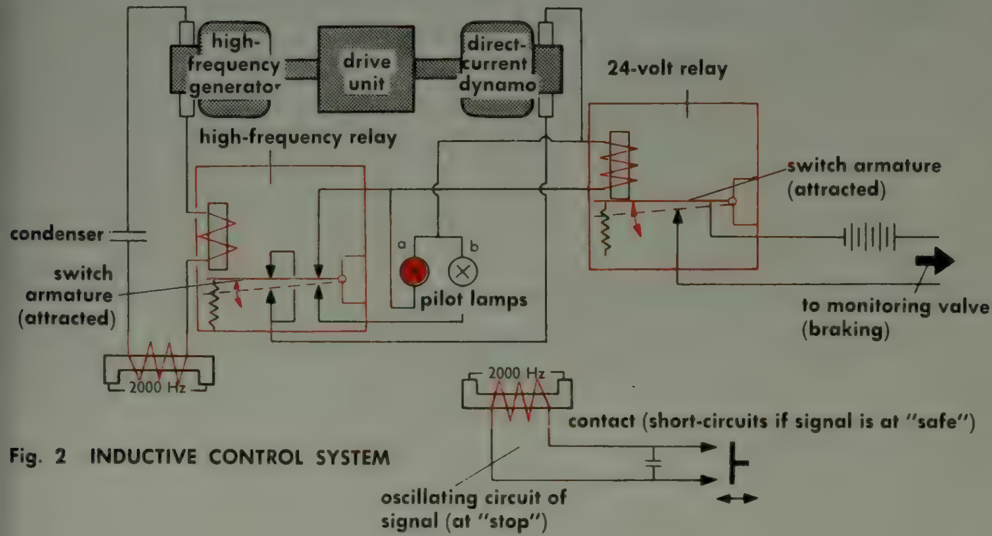
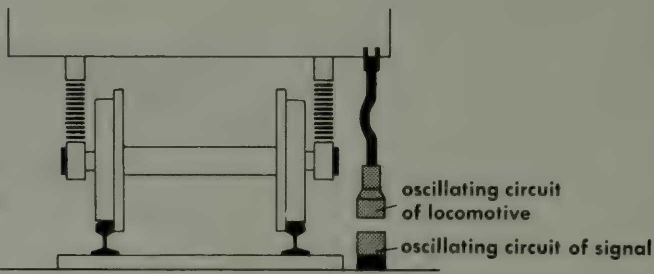


Fig. 2 INDUCTIVE CONTROL SYSTEM

Fig. 3
EXTERNAL EQUIPMENT FOR
INDUCTIVE CONTROL SYSTEM



RAILWAY SWITCHES OR POINTS

The terms "points" and "switches" (the latter more particularly in American practice) are used to denote devices, usually comprising tapered metal blades or tongues, for setting alternative routes of running rails. In the somewhat more general sense of a curved track leading from one track to another, the term "turnout" is also used in this connection. The commonest form of switch is the split switch (Fig. 1) in which one rail of the main track and the inner rail of the turnout are continuous. There are various other types of railway switch, sometimes embodying a combination of two split switches (double turnout) (Figs. 2 and 3). Where two tracks simply cross each other without provision for trains being routed from one track to another, the term "crossing" is usually employed. Fig. 5 shows a common type of crossing (diamond crossing). In some cases, however, more particularly when the two intersecting tracks form a small angle with each other, the crossing may take the form of a so-called crossing switch (also known as "slip points"). In Fig. 4 a device of this kind is illustrated.

The operation of all these devices is similar in principle and can best be explained with reference to the ordinary split switch (Fig. 1): When the straight tongue "a" is in contact with the rail I and the curved tongue "b" is not in contact with the rail II, the switch is set for running straight ahead on the main track. When the switch is set so as to divert a train coming from the left on to the turnout track, the tongue "b" is swung into contact with the rail II and the tongue "a" is now no longer in contact with the rail I. The point of intersection of the inner rails is called the "frog" of the switch, marked by H in Fig. 1. It is usually in the form of a V-shaped unit. As a safeguard against derailment the rails opposite the frog are provided with guard rails (e), and the frog itself is assisted initially (i.e., at its point) by wing rails (f) to carry the weight of the wheels passing over it.

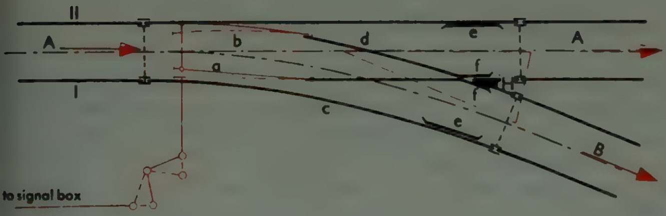


Fig. 1 SPLIT SWITCH

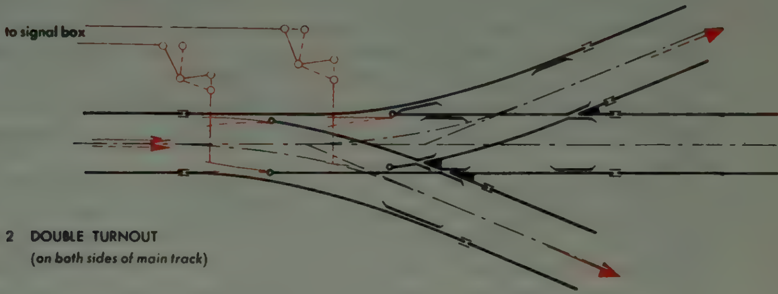


Fig. 2 DOUBLE TURNOUT
(on both sides of main track)

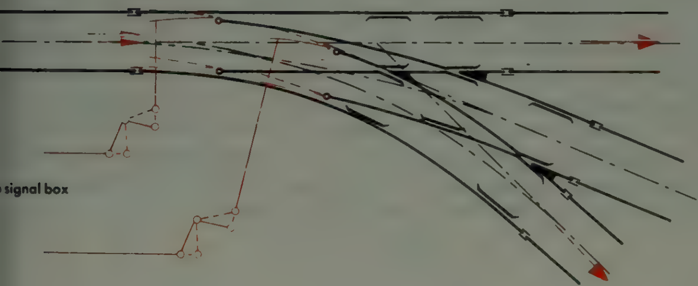
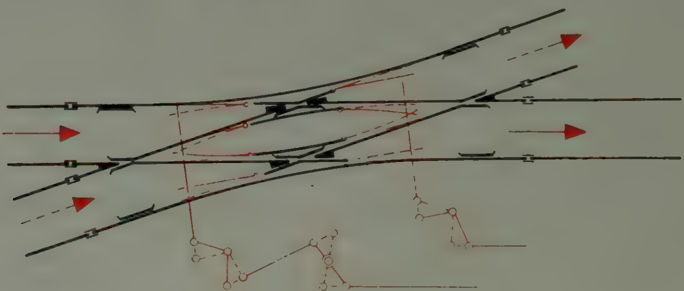
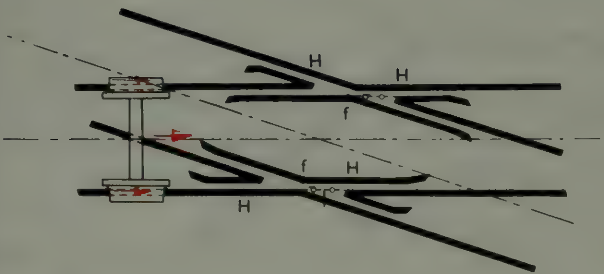


Fig. 3 DOUBLE TURNOUT
(on one side of main track)



ig. 4 CROSSING SWITCH



. 5 DIAMOND CROSSING

The switches (points) and signals on a section of railway track are worked by controls accommodated in a signal box (or tower). On lines carrying scheduled train services these controls are operated in accordance with predetermined timetables.

The principle of mechanical switch and signal operation is illustrated in Fig. 1. In the signal box are pulleys, each connected to an operating lever. When the signalman moves the lever, the pulley is rotated a certain amount, depending on the desired switch position. A catch secures the lever in position. A steel wire rope passes round the pulley. The end of this rope is attached to the switch actuating mechanism or to the operating wheel of the main signal. The tensioning device keeps the wire rope constantly taut. Now when the signalman swings the operating lever, the actuating wheel of the switch mechanism is rotated through a certain angle by the wire rope, so that the actuating lever (connected to the wheel) is likewise swung about its pivot and thus shifts the tongues of the switch to the desired position. The main signal is similarly worked, the arm of the signal being moved by a rod attached to the actuating wheel.

Fig. 2 represents the track layout of a four-track railway station. Two trains are standing on tracks 3 and 4 respectively. The train arriving on track 4 must cross the tracks 2 and 3 and continue its journey on track 1. The switches 1 and 4 and the crossing switches 3 and 2 must be appropriately set. To establish the route for the train approaching on track 4, the signalman must accordingly set the relevant switches, crossing switches and signals. When this has been done, he operates a route lever. This lever, which is associated only with the route marked in red in Fig. 2, can be operated only if all the switches and crossings for this route have first all been correctly set. This dependence of the route lever upon the switch settings is achieved by a system of interlocking controls. A separate route lever is provided for each route. None of these levers can be operated unless all the signals (1a–4a and 1e–4e) are set at "stop". When the route has been established, all the signals except 4a are locked in the "stop" position. When the route is cleared by the station inspector, the signalman shifts the signal 4a to "go". The train on track 4 can then proceed.

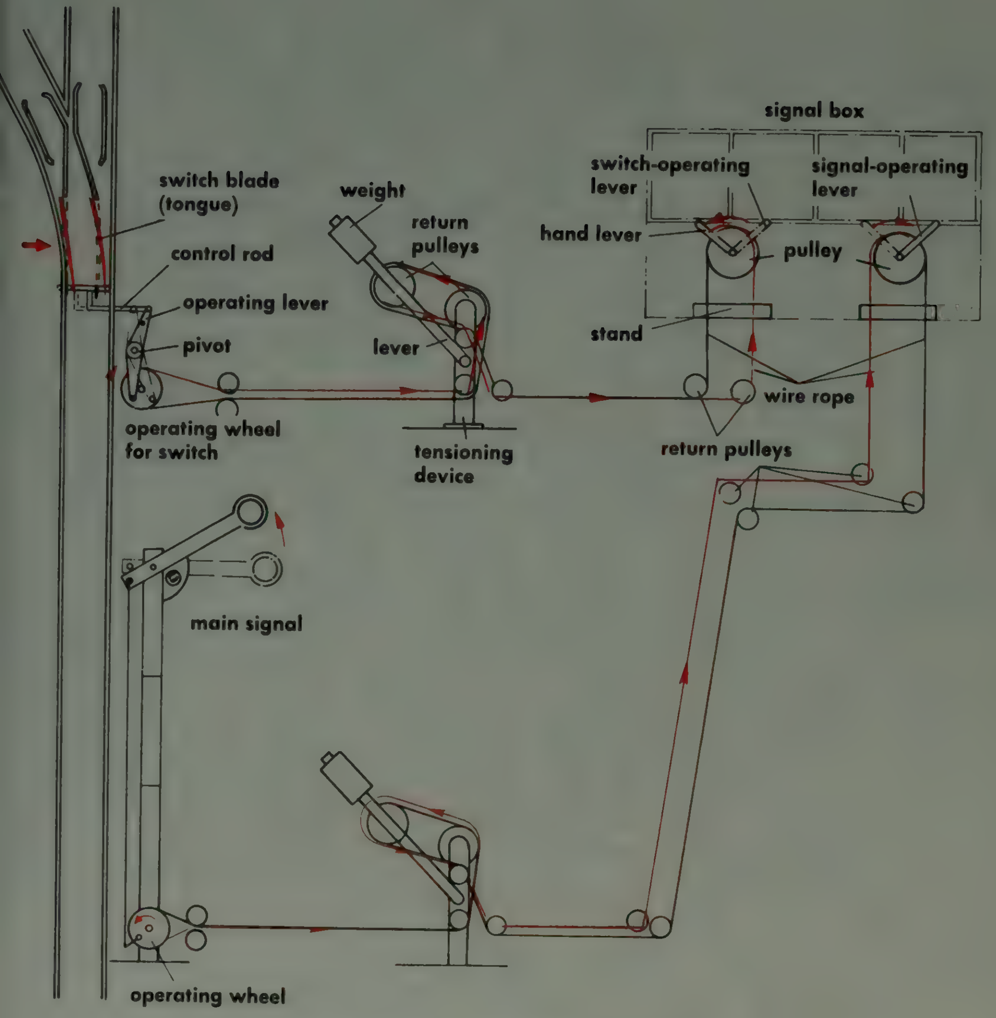


Fig. 1 SWITCH AND SIGNAL OPERATING SYSTEM

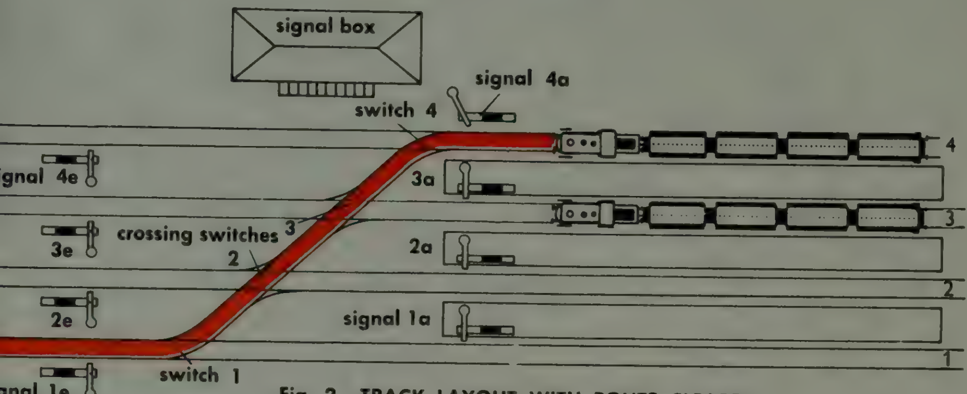


Fig. 2 TRACK LAYOUT WITH ROUTE CLEARED

Railway signal-box: mechanical system
Photo CFF


16
15b

Operating a mechanical signal box (see page 264) involves the exertion of considerable physical force by the signalman. For this reason in modern installations the points and signals are worked by small electric motors. On a control desk in the signal box each switch (set of points) has its own control key (electric switch). Also, there are similar keys for controlling the signals. The signal box also contains, at eye level, an illuminated diagram showing the track layout and all the switches, crossings and signals of the section of railway line concerned. The switch and signal positions are indicated on this diagram by means of small coloured lights. A glance at the panel also shows whether any particular track is free or occupied by a train. An illuminated diagram of this kind greatly facilitates the signalman's task. A further development, providing even greater convenience and sureness of operation, is the "track plan" signal box. In this arrangement the signalman's control desk itself is laid out as a track plan showing all the signals and switches. Each of these is provided with a key or push-button by means of which the corresponding signal or switch on the track can be operated. The tracks themselves are represented on the control desk by small illuminated compartments (Fig. 2). Unoccupied tracks and track sections which at any particular moment are not in use as a train route remain dark, i.e., not lighted up. The switches in the track plan are additionally marked by yellow lamps installed in slots. When these lamps light up, they indicate in which position the switch has been set (Figs. 1a and 1b). In Fig. 1c the switch is set at "straight ahead", but is occupied by a vehicle (compartment lighted red) and is "closed" to all other traffic. In Fig. 1d the switch is set at "straight ahead" and the route is now clear (indicated by a second yellow-lighted slot to the left of the switch). In Fig. 1e the main signal is set at "safe" (green light). Fig. 1f shows the signal at "stop".

A track plan corresponding to the layout in Fig. 2 controlled by the mechanical system described on page 264 is represented in Fig. 2a. Tracks 3 and 4 are each occupied by a train; the corresponding slots in the track plan of the control desk are accordingly lighted red. All the signals are set at "stop". The switches 13–16 are all set at "straight ahead". Now the station inspector gives instructions to clear a route from track 4 to track 1, for example. This instruction is brought to the signalman's attention by the yellow lighting-up of the (hitherto unlighted) slots along the required route. The signalman thereupon works the push-buttons 8 and 17. The winking of the yellow indicator lights of the switches 13–16 informs him that the switches are moving in to the desired positions. The winking light in the slots changes to a steadfast light. The switch positions shown in Fig. 2a change to those in Fig. 2b: the selected route is then indicated by steadfast yellow lights along its entire length. Now the signal 8 changes from red to green, and the train waiting on track 4 can proceed. All the switches on this route are now kept locked by an electric circuit and their positions cannot be altered. Also, the signals 5, 6 and 7 are set at "stop". The train on track 3 therefore cannot move ahead. When the train on track 4 travels along the route thus cleared for it, the yellow-lighted slots change to red, section by section, according to the train's position. Behind the train the red lights in the slots change back to yellow and then soon go out. The route is thus cancelled again and the signal 8 reverts to red. All the switches remain in the positions at which they have been set, however. The position of the train on the track plan is automatically indicated by the train itself.

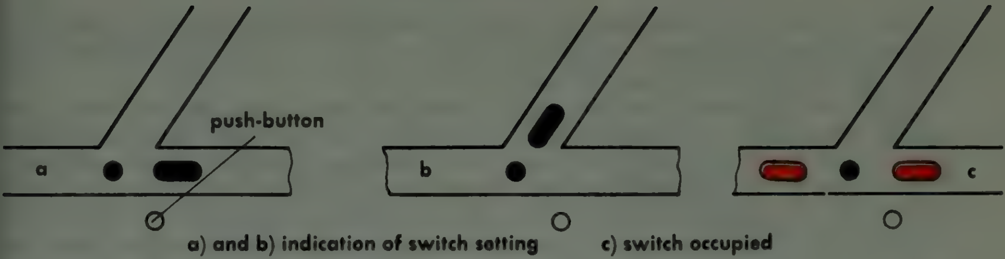
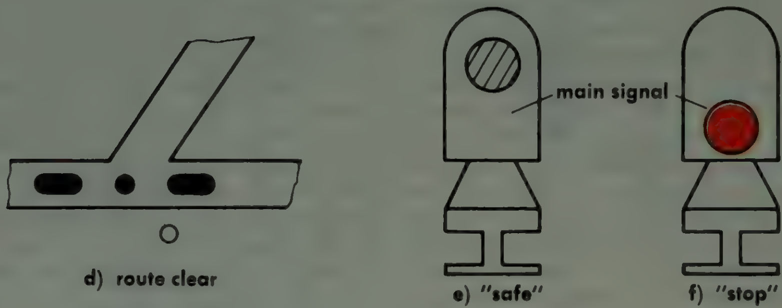


Fig. 1 SWITCH AND SIGNAL SETTINGS



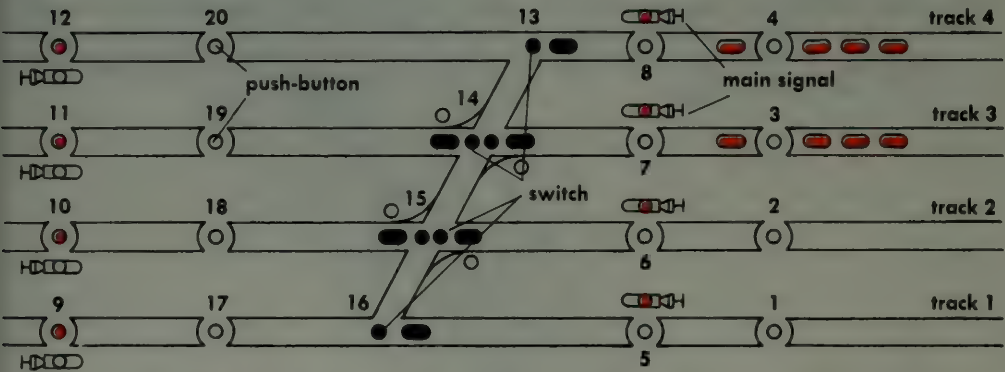
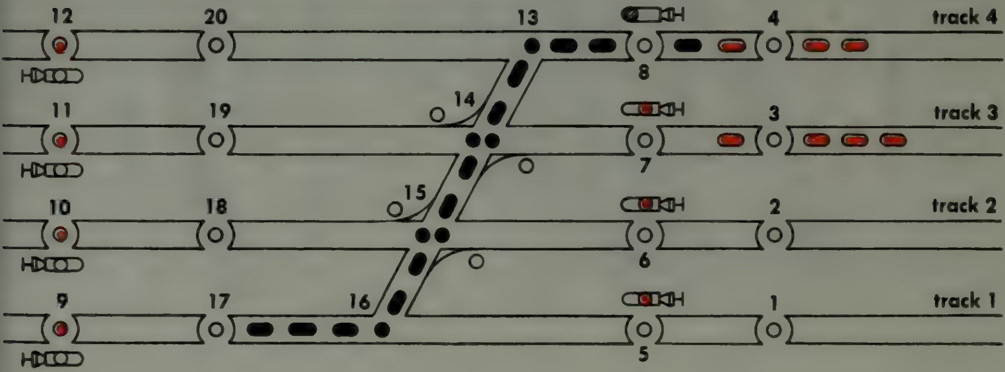


Fig. 2a TRACK PLAN WITH TWO TRACKS OCCUPIED

Fig. 2b TRACK PLAN WITH ROUTE CLEARED



An escalator is a continuously acting passenger conveying device which has about ten times the hourly handling capacity of a lift (p. 272, vol. I). The escalator conveys passengers at a speed of 8–12 ft./sec.

Each individual step of an escalator is constructed as a carriage provided with four small wheels (Figs. 1 and 3). The top and the bottom pair of wheels each run on rails; the rails of the top pair are set farther outwards than those of the bottom pair. On the upward journey of the step the two rails are situated in the same plane, but a short distance before the top reversal point and a short distance past the bottom reversal point the rails are so displaced in relation to one another that the inner rail is below the outer rail. As a result of this arrangement the steps gradually merge into a flat horizontal surface at the top of the escalator, enabling the passengers to step off easily. Similarly, convenient stepping-on is ensured at the bottom of the upward-moving escalator. The moving steps perform the return journey on the rails which are continued on the underside of the escalator (Fig. 2). Each step is attached to two endless chains which run on sprocket wheels at the top and bottom of the escalator; the top sprocket wheels are driven by an electric motor, whereby the steps can be made to travel upwards or downwards. A ratchet wheel is mounted on the drive shaft. If a chain fractures, a pawl engages with the teeth of the ratchet wheel and stops the escalator. The motor is automatically switched off at the same time.

The moving handrails consist of continuous belts made of rubber and canvas plies incorporated in it. They are driven by the escalator drive shaft through a gear wheel and chain system. Half-way along the escalator is a tensioning wheel which is pressed against the return strand of the handrail belt and is thus kept taut, so that it cannot slip off the top and bottom return pulleys. Before and after the tensioning wheel the belt is deflected through an angle of 180°, so as to prevent the belt surface becoming roughened by wear. The guide roller keeps the belt accurately aligned with the bottom return pulley.

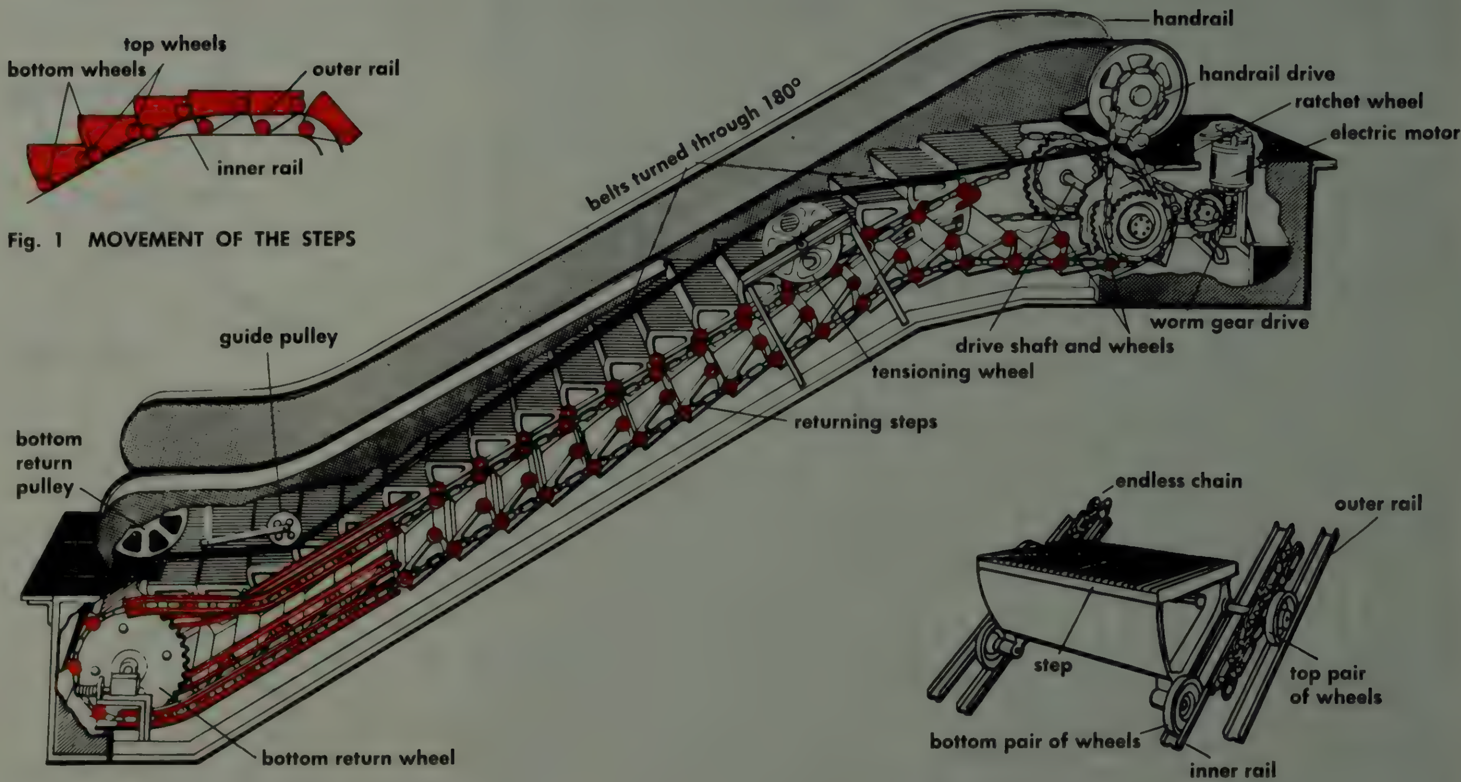


Fig. 1 MOVEMENT OF THE STEPS

Fig. 2 SECTION THROUGH ESCALATOR

Fig. 3 STEP OF ESCALATOR

According to Archimedes' principle, a body which is wholly or partly immersed in a fluid undergoes a loss in weight equal to the weight of fluid which it displaces. An aluminium cube with sides 1 ft. in length weighs about 168 lb. (Fig. 1a). A cubic foot of water weighs about 62 lb. If the aluminium cube is immersed in water (Fig. 1b), its weight has apparently decreased to 106 lb. This is because the cube displaces a cubic foot of water and thereby undergoes a loss in weight equal to the weight of this displaced water. The upward force due to *buoyancy* in this case is equal to 62 lb. and acts at the centre of gravity of the displaced volume of water. If a body, on being totally immersed in a fluid, would displace a volume of fluid whose weight is greater than that of the body concerned, then that body will float on the fluid. Floating merely means that the body sinks into the fluid to such a depth that the displaced volume of fluid weighs exactly as much as the whole floating body. In that case the upward force (buoyancy), which is equal to the weight of the displaced fluid, is in equilibrium with the weight of the body. A 1 ft. wooden cube weighs about 50 lb. It will float in water; the submerged part of the cube displaces a volume of water weighing 50 lb., so that the upward force is 50 lb. and thus counter-balances the weight of the cube (Fig. 2). Hence the displacement of a floating object is equal to its weight.

This is the elementary principle of floating. However, a ship must additionally have stability, i.e., it must be able to right itself after being swung to an inclined position by an external force such as wind pressure. A ship is said to "heel" when it leans over to port or starboard (Fig. 3a); the term "trim" refers to the longitudinal position of a ship in relation to the waterline: the ship is said to be trimmed by the head (as in Fig. 3b) or by the stern, according as the head or the stern lies deeper down in the water. Stability is especially important with regard to the danger of capsising. Fig. 4a shows the ship in its normal position. Its weight can be conceived as a downward force acting at its centre of gravity S. The counterbalancing upward force acts at the centre of buoyancy W, which is the centre of gravity of the displaced volume of water. Normally the points S and W are located on the same vertical line. When the ship heels over (Figs 4b and 4c), the centre of buoyancy shifts to a different position (marked W^1), and the upward force acting here strives to rotate the ship around its centre of gravity S. The intersection M of the line of action of the upward force A with the ship's axis of symmetry is called the metacentre. If the metacentre is located above the centre of gravity S of the ship (as in Fig. 4b), the ship will return to its normal upright position; it is said to be in stable equilibrium. On the other hand, if the metacentre is below the centre of gravity S (Fig. 4c), the ship is in unstable equilibrium and will capsise when it heels over.

Archimedes' principle

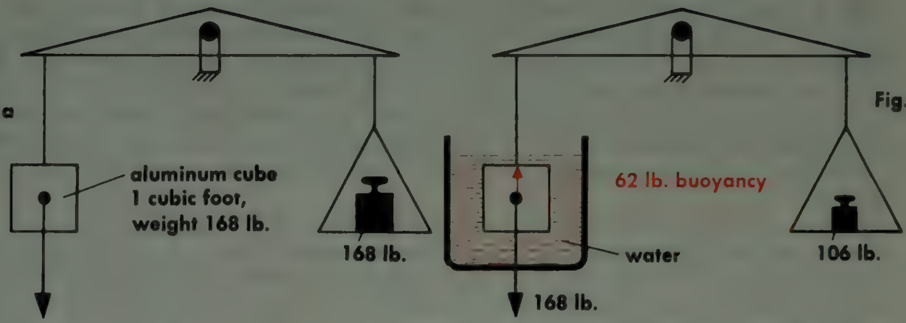


Fig. 1a

Fig. 1b

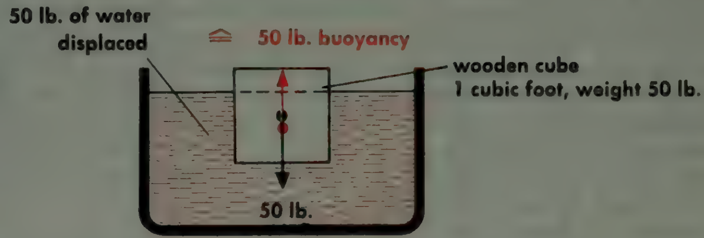
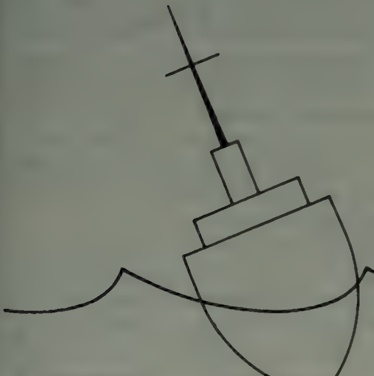


Fig. 2 FLOATING



g. 3a HEELING

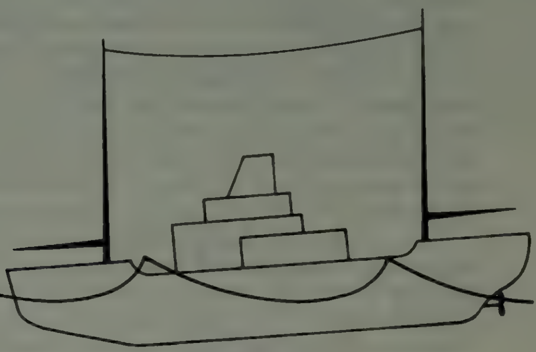


Fig. 3b TRIM

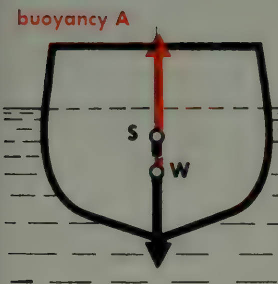


Fig. 4a

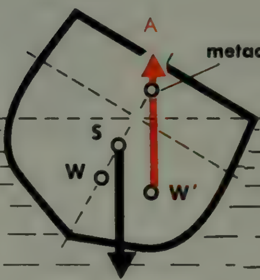


Fig. 4b

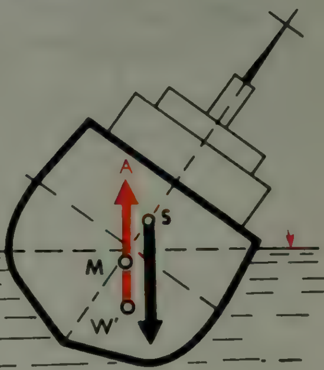


Fig. 4c

The two main types of sail are square sails and fore-and-aft sails. The former are set in a position across the longitudinal axis of the ship, whereas the latter are set along the axis. The sailing ships of olden days were mostly square-rigged. An example of fore-and-aft rig is provided by a typical sailing yacht as illustrated in Fig. 1. This boat has a mainsail, a foresail and a jib. The tall triangular mainsail has better wind-catching efficiency than a mainsail of equal area attached to a gaff fitted to the top part of the (generally lower) mast. At the base the mainsail is attached to the boom which can swing about the lower part of the mast in either direction in relation to the longitudinal axis of the boat.

In the early days of sailing, ships were only able to sail before the wind (Fig. 2a) or with wind on the quarter (Fig. 2b). If they encountered half wind (Fig. 2c) or head wind (contrary wind) (Fig. 2d), they had to wait for the wind to turn. However, as a result of gradual improvement of the design of the hull, rigging and shape of the sails, it became possible to sail a ship against the wind by a procedure known as "tacking" or "beating up against the wind", in which the ship follows a zig-zag path (Fig. 3). The precise functioning of a sail is a somewhat complex phenomenon. For one thing, it acts as a straightforward wind-catching area. Furthermore, because of its curvature under the pressure of the wind, the sail develops aerodynamic behaviour comparable to that of an aircraft wing (cf. page 284), which is associated with a forward suction effect acting on the outside (convex side) of the sail. Thirdly, there is a kind of jet propulsion effect (cf. page 292) produced by the air streaming between foresail and mainsail. It is for this latter reason that racing yachts have foresails extending backward some distance past the mast.

In Fig. 3 the arrow W represents the constant wind force. Of this force only the component S, at right angles to the sail, is effective in propelling the boat. This force S tries to push the boat in the direction of the arrow S. However, the boat encounters a high resistance from the water (pressing against hull and keep) in this direction. Hence, only the longitudinal component V of the force S acts as a forward-propelling force on the boat. On reaching the point C, the helmsman swings the boat round towards D and at the same time swings the sail so that its surface is once again approximately bisecting the angle between the direction of the wind and the longitudinal axis of the boat. The boat is now propelled in the direction C–D. At D the boat and the sail are once again swung round, and so on.

Most yachts and larger vessels have lead-ballasted fixed keels (Fig. 1), but some smaller boats have movable keels (Fig. 5). The ballast in a fixed keel is so designed in relation to the sails that, in conjunction with the lateral resistance of the water, it normally more than counteracts the overturning effect of the wind pressure (Fig. 4a). The movable keel (or centre-board) of a small boat such as a dinghy has no ballast, and to avoid capsizing, the crew must lean far out to counterbalance the wind pressure on the sail (Fig. 4b).

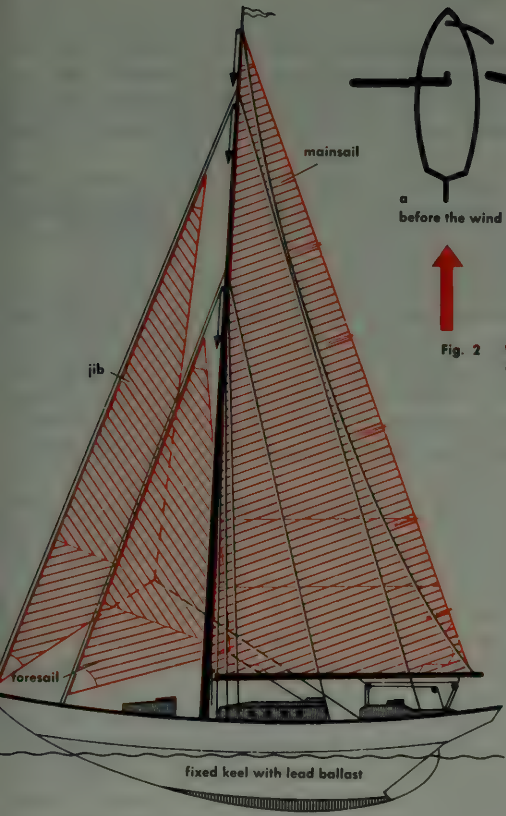


Fig. 1 YACHT WITH FORE-AND-AFT RIG

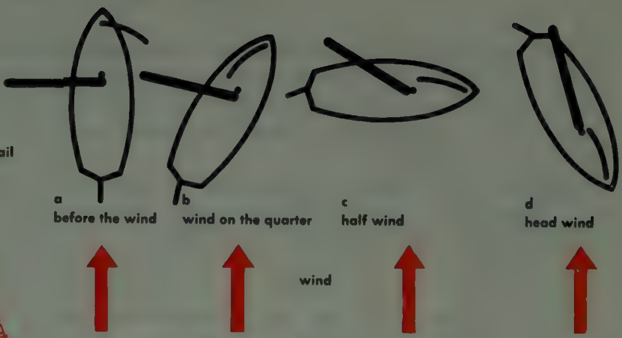


Fig. 2 WIND DIRECTIONS IN RELATION TO BOAT

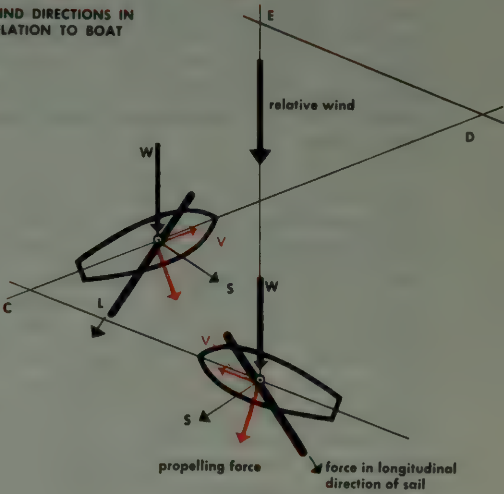
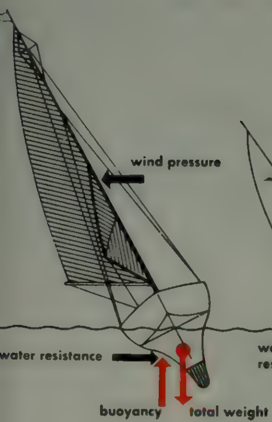
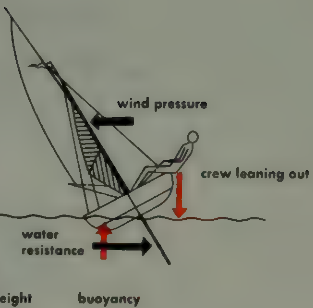


Fig. 3 TACKING





ig. 4a FORCES ACTING ON BOAT WITH FIXED KEEL

Fig. 4b FORCES ACTING ON BOAT WITH MOVABLE KEEL

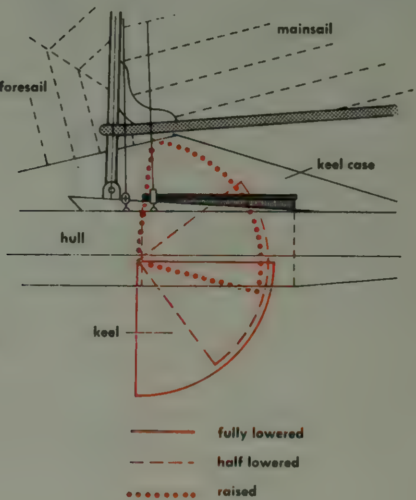


Fig. 5 MOVABLE KEEL

Like all propellers, the function of the ship's propeller (technically called "marine screw propeller") is to produce the thrust to drive the ship by giving momentum to the water it displaces in an astern direction. The propeller, as it were, pushes the water backwards and this in turn develops a reaction force which thrusts the ship forward (cf. page 312) (Fig. 2).

The principle of the screw motion of a propeller is illustrated in Fig. 1. The tip of each revolving blade describes a helical curve as the propeller moves forward in the direction of movement of the ship. The pitch is the longitudinal distance corresponding to one complete revolution of the propeller (Fig. 1).

The efficiency of propulsion, i.e., the proportion of the engine power output that is utilised for propelling the ship through the water, is determined by the difference between the approach velocity of the water ahead of the propeller (which velocity is equal to the speed of the ship) and the velocity of the water displaced astern of the propeller at each revolution. When a ship starts up its engines and begins to move, small quantities of water are given a large sternward acceleration by the propeller; when the ship is under way, large quantities of water are given a relatively small acceleration.

In a solid unyielding medium, each revolution of the propeller would cause the ship to travel a distance equal to the pitch of the propeller. In reality, water is a yielding substance which gives way a little under the pressure of the propeller blades. As a result of this, the actual forward motion achieved at each revolution is only about 60–70% of the pitch. The difference in relation to 100% is known as the slip of the propeller. On the suction side of the propeller a negative pressure is produced, which is greater according as the angle of incidence of the blades is larger and their speed of rotation is higher. If the negative pressure is too great, the flow of water round the propeller blades is disrupted and bubbles filled with water vapour are formed (this phenomenon is known as cavitation); for this reason high-speed screw propellers are given very wide and flat blades with low angles of incidence (Fig. 3). On the other hand, a low-speed propeller has narrow blades with a high angle of incidence (Fig. 4). Large slow-running propellers (large masses of water, low acceleration) are preferred for ocean liners, whereas small inland navigation craft are usually provided with small fast-running propellers (small masses of water to which a high acceleration is imparted).

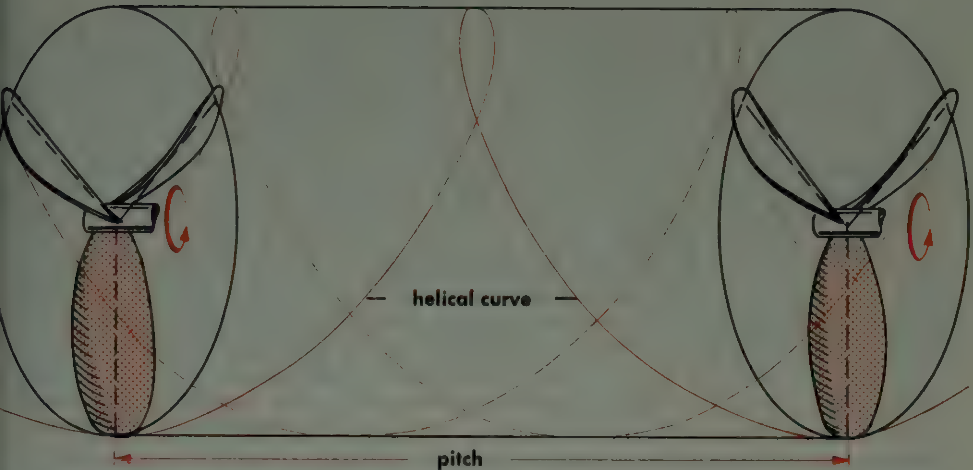


Fig. 1 PRINCIPLE OF SCREW MOTION

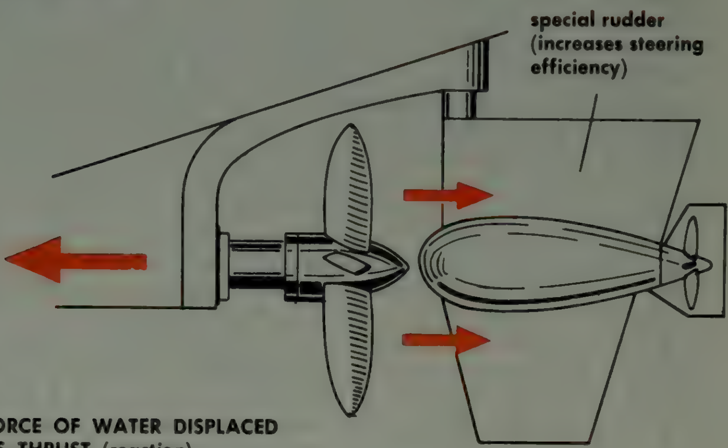


Fig. 2 ACCELERATIVE FORCE OF WATER DISPLACED (*action*) PROVIDES THRUST (*reaction*)

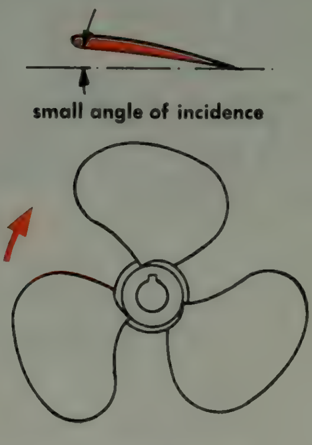


Fig. 3 HIGH-SPEED PROPELLER (*wide, flat blades*)

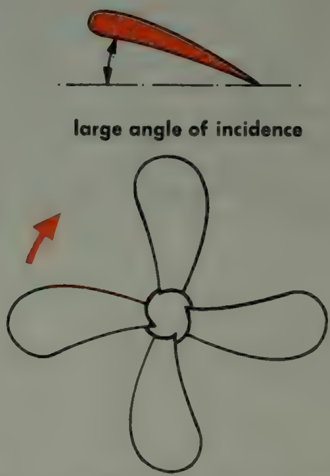


Fig. 4 LOW-SPEED PROPELLER (*narrow blades strongly curved in section*)

By navigation is understood the science of determining the position and course of ships and aircraft (for air navigation cf. page 300 *et seq.*). In marine navigation the ship's position is usually determined as the intersection of two or more position lines. The latter are determined by the taking of bearings on terrestrial or celestial fixed points. Terrestrial fixed points are landmarks, buoys, beacons, lighthouses, etc. Fig. 1 shows how a navigator takes his bearings on two such points. By means of his sextant (cf. page 280) he measures the angle α between the two points A and B which are a known distance apart. Also, he measures the angle β between one of the position lines and the north direction indicated by the ship's compass. As the geographical latitude and longitude of the points A and B, as well as the distance between them, are known, one side and two angles in each of the triangles ABS (or ABS') are known. The lengths of the position lines and the geographical co-ordinates of the position S (or S'). In the open sea, out of sight of land, the navigator must take bearings on a heavenly body or make use of radio direction-finding (see page 302 *et seq.*). In celestial (or astronomical) navigation the navigator, using an instrument called a sextant, sights his instrument, for example, at a star and measures its culmination altitude, i.e., the angle between the line of sight and the horizon (Fig. 2). The position of the star in the sky at that particular time can be looked up in tables (nautical almanacs). The ship's chronometer, which is a very accurate timepiece, gives the precise time at which the observation is made. From these data the position of the ship can be determined. The successive positions determined in this way are plotted on a nautical chart. Among other information, charts also show the depths of water. An important feature of navigation in coastal waters is determining the depth or taking soundings. Formerly a lead line was always used for the purpose; a more modern device for measuring depths is the echo-sounder (see page 280).

The couse plotted between two points A and B can either be an orthodrome, i.e., the shortest distance between these points measured along a great circle of the globe, or a loxodrome (Fig. 3a). The latter course is characterised by the fact that it intersects the circles of longitude always at a constant angle. On a map drawn to Mercator's projection the loxodrome therefore appears as a straight line (Fig. 3b). The loxodrome is nevertheless longer than the orthodrome.

Radio direction-finding signals and bearings on lighthouses provide the navigator with reference points. The exact geographical location and the characteristics of each lighthouse can be looked up in tables. For example, Fig. 4 shows the characteristics of a revolving light which periodically emits short and long flashes.

For keeping the ship on course, regulating its speed, and manoeuvring in harbours and in perilous situations a very important piece of equipment is the telegraph (Fig. 5). It is used for transmitting orders from the bridge to the engine-room. When the lever is moved to a particular position of the dial, a sprocket moves a chain to whose ends wires are attached which actuate a similar dial in the engine room. In the more modern form of the telegraph the signal is transmitted electrically. For example, when the lever of the bridge telegraph is moved to "half ahead", the pointer on the engine-room telegraph also moves to "half ahead", and a bell rings to attract attention. The engineer then replies by moving the lever of his telegraph to "half ahead", and by doing so operates the return pointer on the bridge telegraph, which rings a bell, thus indicating that the order has been understood.

(Continued)

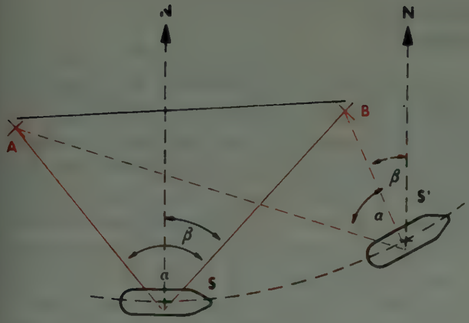


Fig. 1 TAKING BEARINGS ON TWO FIXED POINTS

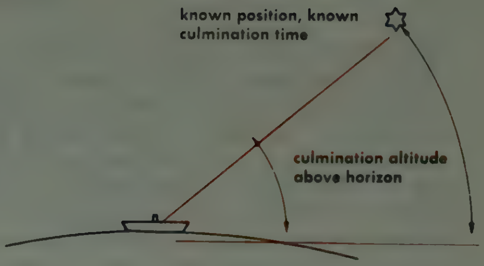


Fig. 2 TAKING BEARINGS WITH THE AID OF A STAR

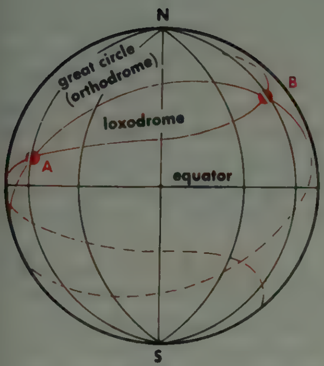


Fig. 3a

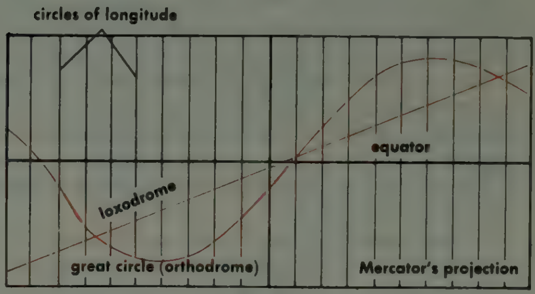


Fig. 3b

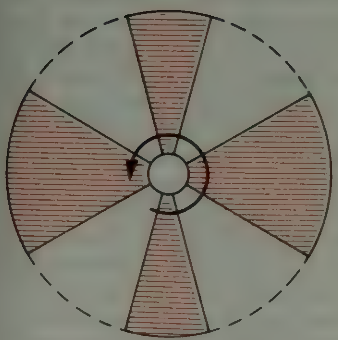


Fig. 4 CHARACTERISTICS OF A REVOLVING LIGHT

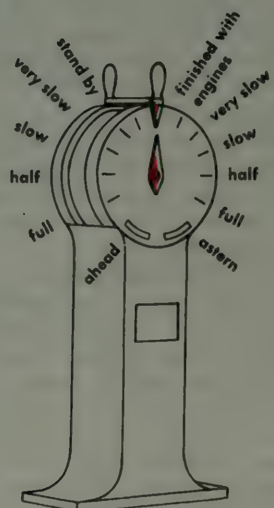


Fig. 5 TELEGRAPH

One of the most important navigation instruments is, of course, the compass (magnetic compass or gyro compass; see page 282). The compass is used for steering and for fixing the ship's position by bearings. The chronometer and sextant have already been mentioned. The *sextant* is used in astronomical navigation for measuring the altitude of a star (or the sun) above the horizon. A simple sextant is shown in Fig. 6. Looking through the hole in the sight vane, the observer sights the instrument at the star, which he sees in the unsilvered half of the mirror. If he wishes to measure, say, the angular distance between this star and another star, the observer rotates an adjustable mirror mounted on a swivel arm until the second star (whose light is reflected from the adjustable mirror and from the half-silvered mirror, as indicated in Fig. 6) is seen to coincide with the first. The angular distance between the two stars can then be read from the graduated scale (which is usually 60°, i.e., one-sixth of a circle, hence the name "sextant"). Most modern sextants are fitted with a small telescope instead of a mere sight hole.

For plotting the course of a ship it is also important to know the speed at which it is sailing. This is measured by means of a device called a *log*. In modern navigation the so-called patent log is mostly used (Fig. 7). A small propeller, called a "rotator", is towed along behind the ship and revolves faster according as the speed is higher. The rotation is transmitted through a flexible shaft to the register fitted with a dial on which the speed (usually in nautical miles) is indicated. A more recent development is the pressure type speed recorder based on Prandtl's Pitot tube (Fig. 8). It comprises two concentric tubes, the outer one being provided with lateral holes. When this device moves through the water, a difference in pressure develops which is measured and which is proportional to the square of the vessel's speed. A similar device is used in air navigation (see page 302) and for measuring flow velocities in aerodynamics and hydrodynamics.

As the water depths are marked on nautical charts, a further navigational aid is provided by soundings, i.e., systematic measurements of the depth of the water in which the ship is sailing. The traditional device for the purpose is the lead-line. Patent sounding machines are based on the fact that the pressure of the water on an immersed body increases with the depth to which it is immersed. A more modern device is the *echo sounder*, one form of which is shown diagrammatically in Fig. 10. In the bottom of the vessel's hull a transmitting oscillator and a receiving oscillator are so mounted that the latter picks up the echo reflected back from the sea bed. A rotating contact causes a condenser to discharge through the electromagnetic transmitting oscillator, so that a sound impulse is transmitted. The time measuring equipment comprises a neon lamp rotating in front of a timing scale. When the lamp passes the zero position, the sound impulse is transmitted. The time it takes for this impulse to reach the sea bed and return to the receiving oscillator (which is essentially a microphone for picking up the sound) is marked, or read, on the scale by a flash emitted by the neon lamp, which has meanwhile rotated past the zero position. The current from the receiving oscillator has to be amplified in order to cause the lamp to light up. Echo sounders usually operate with ultrasonic frequencies (20,000 cycles/sec.), but sounders operating with an audible sound frequency (3600 cycles/sec.) are used for deep-sea soundings. The principle of the transmitting oscillator is illustrated in Fig. 9. The coil is energised by an alternating current with a frequency of 3600 cycles/sec., so that the laminated armature, attached to the diaphragm, is alternately attracted and released. The diaphragm emits sound waves of this same frequency. Other types of transmitter make use of the principle of magnetostriction or the piezo-electric effect (cf. p. 236, vol. I).

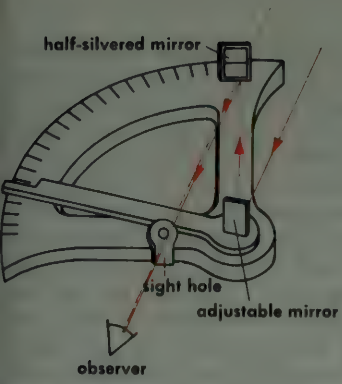


Fig. 6 SEXTANT

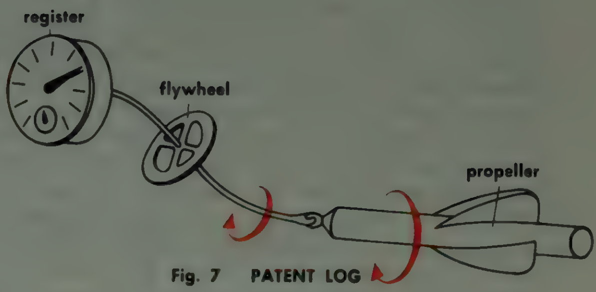


Fig. 7 PATENT LOG

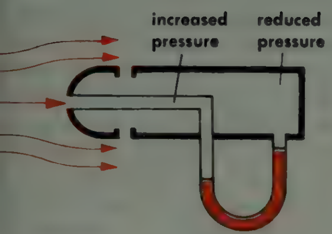


Fig. 8 PITOT TUBE

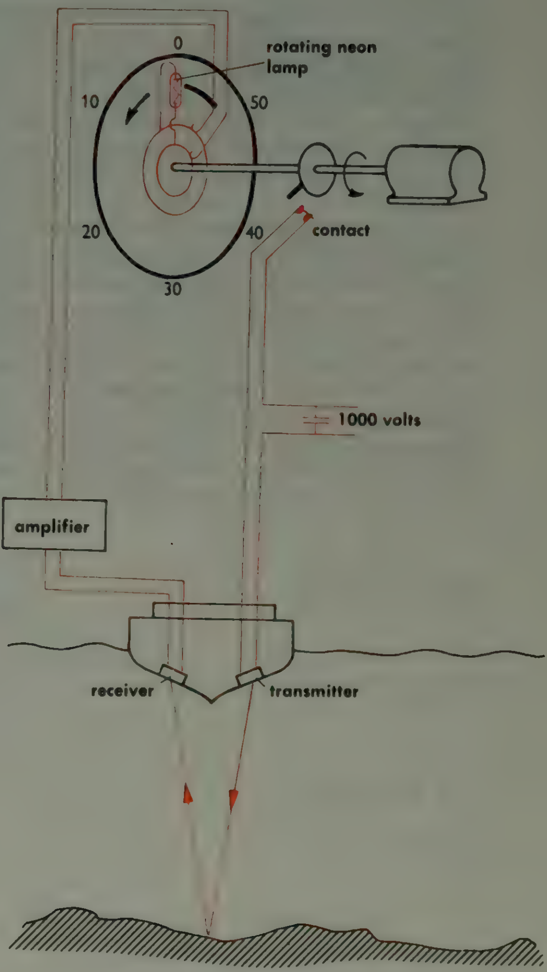


Fig. 10 ECHO SOUNDING EQUIPMENT

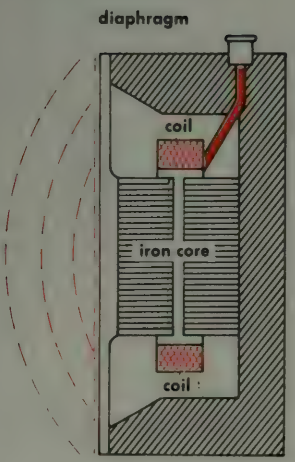


Fig. 9 ELECTROMAGNETIC TRANSMITTING OSCILLATOR

The property of natural magnetism possessed by certain iron ores can be explained with reference to a simple conception of the structure of the atom (Fig. 1). A single electron orbiting around the nucleus acts as an electric current and produces an electric field, i.e., it behaves like a tiny magnet. Iron can be conceived as consisting of a multitude of such elementary magnets which are aligned parallel to one another in groups. Externally the magnetism of these groups cancels out (Fig. 2a). However, if all the groups are aligned in the same direction as the result of external influences, this alignment will, in the main, be retained (Fig. 2b). A piece of iron in which this has happened is a natural magnet. The earth as a whole, too, is a natural magnet (Fig. 3). The magnetic axis does not coincide with the axis of rotation, so that the magnetic poles are displaced in relation to the geographical poles. In navigation the earth's magnetic field is utilised for the determining of direction. The instrument used for the purpose is the magnetic compass. It consists of a non-magnetic housing (e.g., made of brass or plastic) in which a magnetic needle, poised on a point, can rotate freely over a compass card (Fig. 4). The needle swings to such a position that one end (its magnetic north pole) points approximately—not exactly, for the reason mentioned above—to the north and, consequently, the other end points to the south. The magnetic needle's deviation from the true north direction is called the "variation". It varies from place to place and is determined by the direction of the horizontal component of the earth's magnetic field (magnetic declination). The vertical lines of force of the earth's field can be detected with the aid of a magnetic needle pivoted about a horizontal axis (Fig. 5; magnetic inclination, inclination needle). If the variation is marked on the compass card, or if the magnetic declination is known from tables, the exact northerly direction (and therefore all the other significant directions) can be determined. The gyro compass functions on a different principle: a rotor which is rotatably mounted at its centre of gravity tends to dispose its axis of rotation parallel to the earth's axis (Fig. 6). For the sake of simplicity, a gyroscope—i.e., a suitably mounted rotor—is assumed to be located at the earth's equator (Fig. 6, position I). As a result of the rotation of the earth, the gyroscope moves to position II. Since it tends to maintain the position of its axis, its centre of gravity will be raised some distance (because of the nature of the gyroscope mounting). The force of gravity then exerts a torque (twisting movement) which strives to swing the gyroscope and its mounting back to its original position (small black arrows). As a result of this the axis of rotation of the gyroscope swings in the direction of this torque, causing the gyroscope to align itself in the north-south direction. The gyro compass is the directional reference most used by modern navigators. Its advantage is that it indicates true instead of magnetic north and is subject to fewer errors. The magnetic compass is installed as a standby.

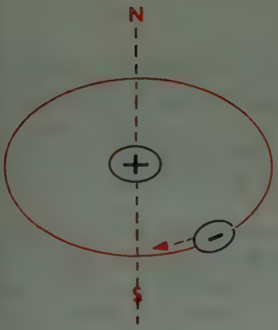


Fig. 1 THE ATOM AS AN ELEMENTARY MAGNET

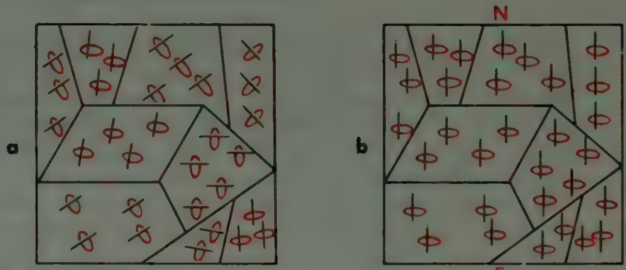


Fig. 2 DOMAINS (*groups of atomic magnets*) IN IRON, SHOWING HOW MAGNETISATION IS EFFECTED

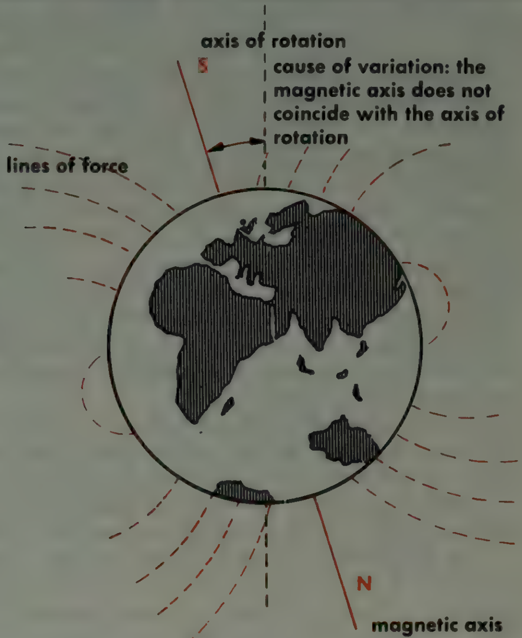


Fig. 3 MAGNETIC FIELD OF THE EARTH

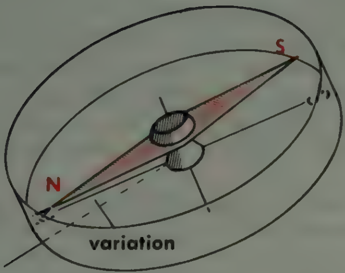


Fig. 4 COMPASS

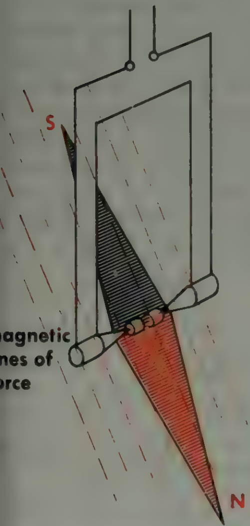


Fig. 5 INCLINATION NEEDLE

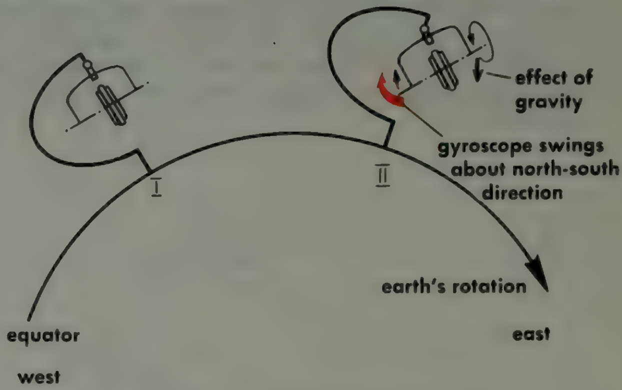


Fig. 6 PRINCIPLE OF THE GYRO COMPASS

As an aircraft is "heavier than air", it needs an upward force to keep it aloft. This force is provided by the "lift" developed by the supporting surfaces (wings) and is directed at right angles to the direction of movement. In addition, the air offers a certain frictional resistance, called "drag" (Fig. 1). By suitable design of the cross-sectional shape of the wing (Fig. 2), the drag can be kept small in relation to the lift. A shape of this kind is known as an airfoil section.

When the wing of an aircraft (Fig. 3) moves forward through the air, the flow of air along the lower surface arrives at the trailing edge before the flow along the upper surface. The lower surface flow attempts to expand around the trailing edge. As a result of this a vortex is formed. The rotation of this vortex accelerates the upper surface flow, so that the length of time required for a particle of air to move from the leading edge to the trailing edge becomes the same for the upper and the lower surface flow. The increased velocity of the upper surface flow eliminates the formation of a vortex by the lower surface air at the trailing edge, and it produces a lower pressure at each point on the upper surface than exists at the corresponding points on the lower surface. It is this difference in pressure that produces the lift. The distribution of lift along the cross-section of a wing is illustrated by the pressure-distribution diagram (Fig. 4). The magnitude of the forces changes with the angle of attack (or angle of incidence), i.e., the angle between the direction of the air flow and the chord line of the wing (Fig. 5). The resultant aerodynamic force acts at the centre of pressure (Fig. 4); its position varies with the angle of attack. The stability of an aircraft is significantly determined by the displacement of the centre of pressure. With increasing angle of attack this point moves forward. When the angle is increased beyond the value that produces maximum lift, "stall" occurs: this results in loss of flying speed and lift and finally loss of control; the air flow detaches itself from the upper surface (Fig. 6).

The airfoil section is so shaped as to present minimum air resistance at the design speed of the aircraft and at the same time provide the necessary amount of lift. Figs. 7a and 7b represent the wing sections of a cargo-carrying aircraft and a faster aircraft respectively. The lift provided by the thick highly curved wing is about half as much again as that of the thinner and flatter wing, but its drag is about twice as high.

The principles of airfoil design are also applicable to the propeller blades. The function of the propeller is to convert the torque developed by the engine into a propulsive thrust to drive the aircraft forward. This thrust is produced by acceleration of the air around the propeller. Since the velocity at each blade section is a function of radius, the blades are twisted to maintain a favourable angle of attack all along the blade. The principal forces and velocities associated with the action of the propeller are shown in Fig. 8. The pitch angle of the propeller blade corresponds to the angle of attack of the wing; it is the angle between the blade chord line and the plane of rotation. A variable-pitch propeller is designed to maintain propeller efficiency as the forward velocity changes. The pitch setting can be changed while the propeller is rotating.

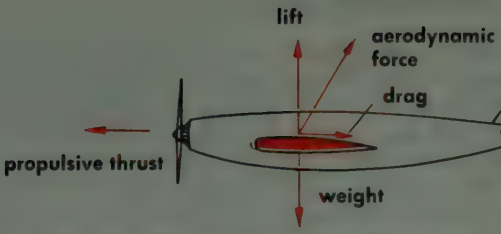


Fig. 1 FORCES ACTING ON AIRCRAFT'S WING

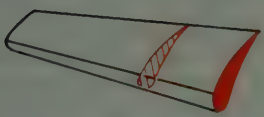


Fig. 2 AIRFOIL SECTION

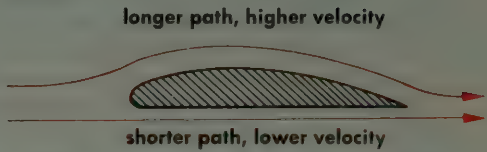
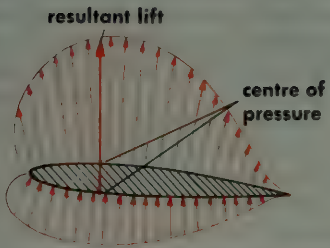


Fig. 3 AIR FLOW AT WING



g. 4 PRESSURE DISTRIBUTION DIAGRAM

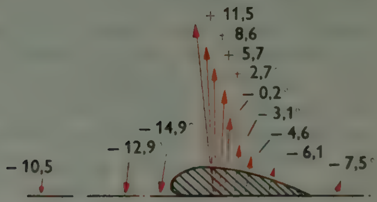
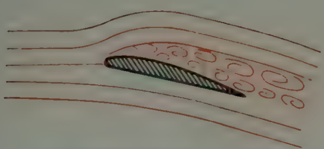


Fig. 5 VARIATION OF FORCES AND POSITION OF CENTRE OF PRESSURE WITH ANGLE OF ATTACK



g. 6 CO-OPERATION OF FORCES ACTING ON PROPELLER AND WING

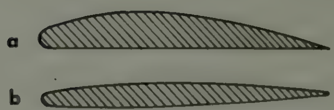


Fig. 7 DIFFERENT AIRFOILS

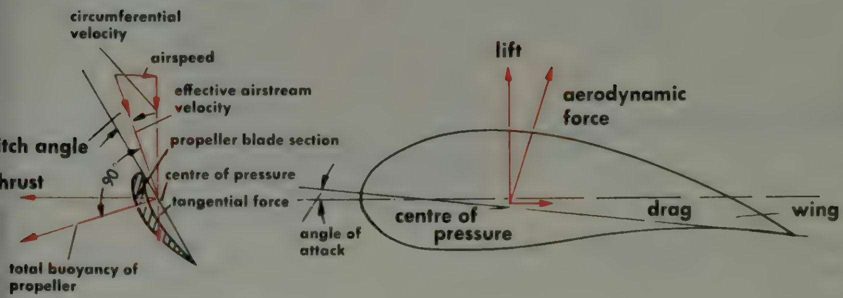


Fig. 8

The speed, or velocity, of sound is the speed at which a disturbance is propagated through a substance (in particular: air). It depends on the elastic properties and density of the substance concerned. In air at normal atmospheric pressure, at a temperature of 15° C, sound travels at a speed of 1120 ft./sec. (340 m/sec.), i.e., about 760 m.p.h. So long as the speed of an aircraft travels at a speed below the speed of sound, the air will be displaced in all directions and behave as an entirely "soft" medium. However, when the aircraft reaches the speed of sound, the air "becomes hard" and presents a very high resistance—the "sound barrier"—which the aircraft must break through or, rather, slice through. On going through the barrier, the aircraft produces a series of elastic waves in the air. As they travel at a slower speed, these waves lag behind the aircraft. The transition from supersonic to subsonic air flow is accompanied by a shock wave.

What happens may be compared to cutting through a material with a knife. (Fig. 1). For example, when a knife cuts a sheet off a roll of paper, a "hissing" noise is heard. These sound waves are produced by the knife blade cutting through the paper. The same thing happens when a body travels through the air at supersonic speed, i.e., faster than sound. The air behaves like a solid substance towards the "knife" cutting through it. The leading part of the body, e.g., a shell fired from a gun (Fig. 2), produces a shock wave (bow wave), while other waves emanate from the rear end (tail wave). Behind the projectile a vacuum is formed, and the adjacent air rushes in to fill this, thereby producing a state of turbulence called a vortex path. To facilitate their passage through the air, projectiles, rockets and supersonic aircraft have pointed noses (Fig. 3). The wings of a supersonic aircraft are relatively much smaller than those of a subsonic aircraft in order to minimise drag. Because of the very high speed, these narrow wings are nevertheless able to develop sufficient lift to keep the aircraft in the air (cf. page 284). The bow wave produced by a supersonic aircraft spreads out in the shape of a cone (Fig. 4). The noise of the aircraft is heard in the zone formed by the intersection of this cone with the ground. The sound pressures recorded in this zone may be as high as 30 lb./ft.^{2}—high enough to shatter windows ("sonic bang").[1] However, if the aircraft flies above about 30,000 ft., such high pressures will not occur.

1. "Sonic boom" in U.S.A.

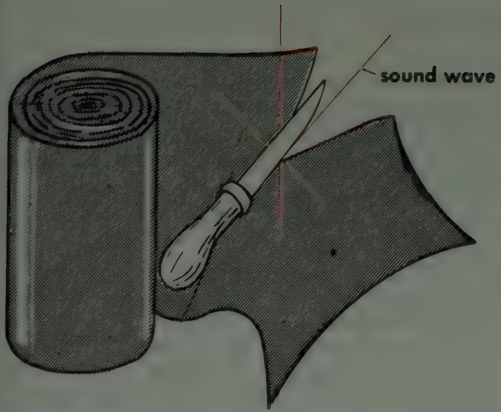


Fig. 1

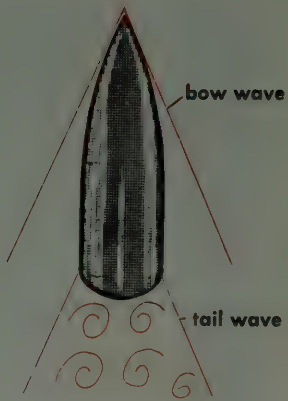


Fig. 2

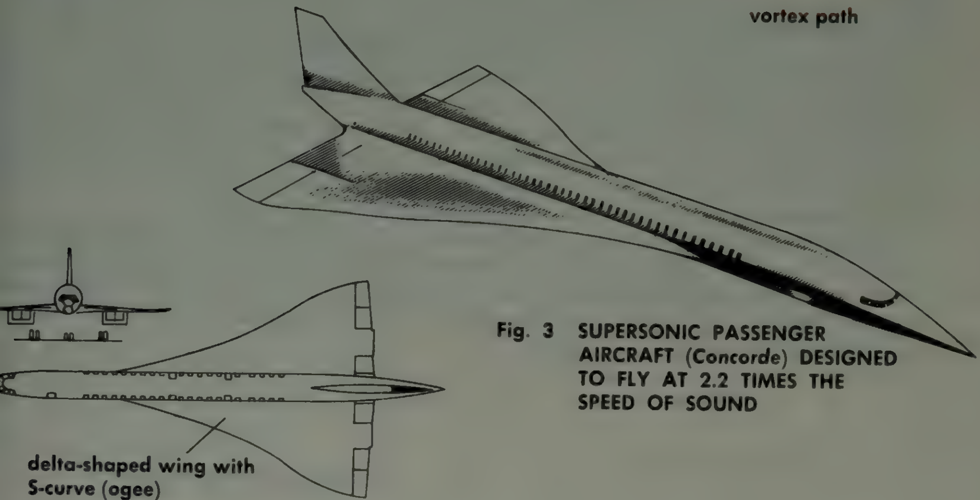


Fig. 3 SUPERSONIC PASSENGER AIRCRAFT (*Concorde*) DESIGNED TO FLY AT 2.2 TIMES THE SPEED OF SOUND

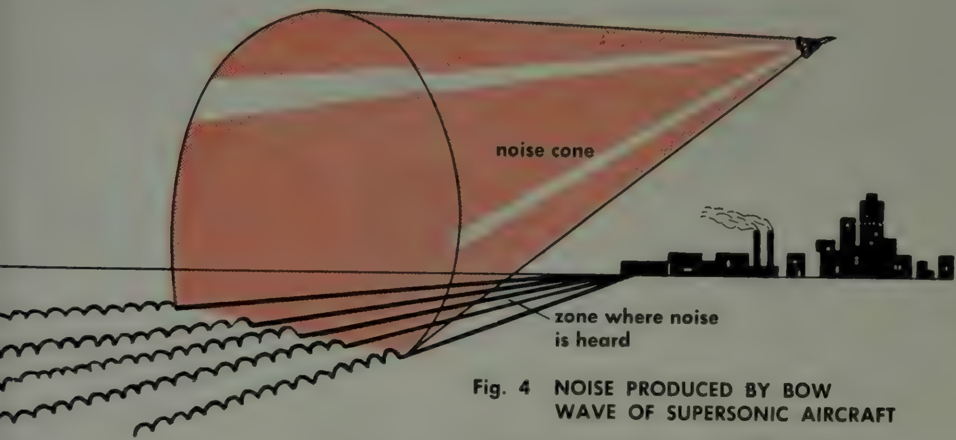


Fig. 4 NOISE PRODUCED BY BOW WAVE OF SUPERSONIC AIRCRAFT

For horizontal flight the weight of an aircraft must be counteracted by the lift (cf. page 284) and the drag overcome by a propulsive thrust acting in the direction of flight. In a powered aircraft (Fig. 1) the thrust is developed by the engine. A glider has no engine. It therefore cannot perform horizontal flight (i.e., at constant height) in perfectly calm weather; under such conditions the glider is merely able to proceed along a downward-sloping path. Only then will the lift counterbalance the weight of the aircraft (or rather the component of the weight perpendicular to the flight path), while the drag is overcome by the forward component of the weight (Fig. 2). The path of the glider (Figs. 3a and 3b) is dependent upon the ratio of drag to lift (gliding angle). The greater the drag, the steeper is the slope of the flight path. The flight performance of a glider is therefore better according as the drag is smaller for a given lift.

The gliding angle determines the duration of the flight under calm weather conditions. However, the attraction of gliding as a sport consists in seeking out and utilising upward air currents which enable the skilled glider pilot to soar higher and higher and thus prolong his flight. This calls for considerable experience and knowledge of weather conditions. For example, birds soaring in the air without flapping their wings are a sure sign of a rising air current, and the experienced pilot will not fail to take advantage of this. A glider, though "heavier than air", can rise if the velocity of the rising air is greater than the velocity of descent of the glider. The rate of climb or descent is indicated by an instrument called a variometer (rate-of-climb indicator, see page 300). Rising air currents develop on hillsides (Fig. 4), over warm areas of land, and under cumulus clouds in the process of formation (thermal air currents). Suitable rising currents also occur along fronts between warm and cold air masses such as are to be found at the edges of thunderstorms.

Launching a glider is usually done by towing the aircraft into the air by an automobile at the end of a long rope (500 ft. or more in length) or by means of a high-speed winch (auto-tow and winch-tow launching technique). Alternatively, airplane tow can be employed, the glider being towed into the air behind a power-driven aircraft and released at the desired altitude.

Fig. 1

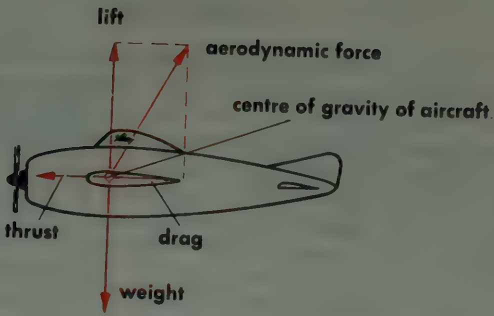


Fig. 2

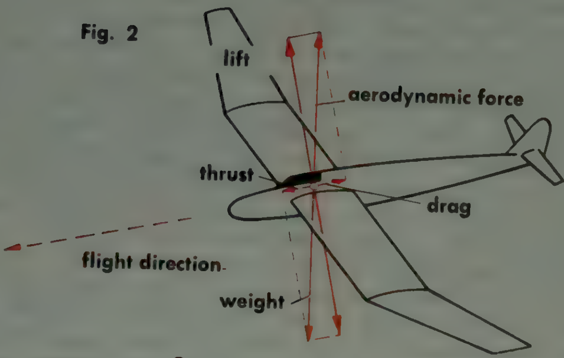
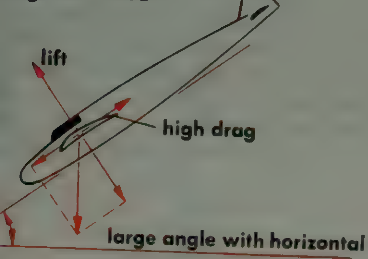


Fig. 3a DIVE



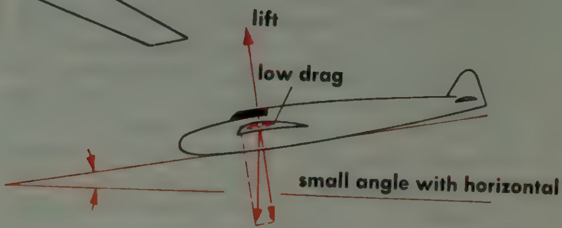
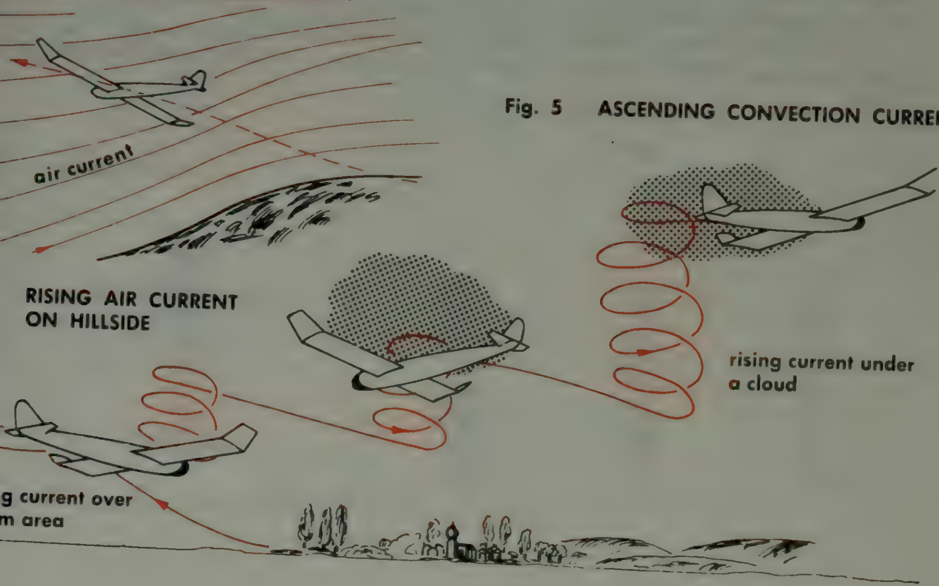


Fig. 3b GLIDE



RISING AIR CURRENT ON HILLSIDE

Fig. 5 ASCENDING CONVECTION CURRENT

A helicopter is equipped with one or more power-driven rotors (horizontal propellers) in lieu of fixed wings. It is able to take off and land vertically, to move in any direction, or to remain stationary in the air. The lift developed by a conventional aircraft wing depends on two factors: the angle of attack of the wing (Fig. 1) and the velocity of the air in relation to the wing. To obtain the necessary lift, the aircraft must have a forward movement (see page 284). In the case of the helicopter the (relative) air velocity is produced by the rotation of the rotor blades: when the angle of attack attains a certain value, the lift overcomes the weight of the aircraft. The aircraft then takes off vertically (Fig. 2). To achieve horizontal flight, the pilot tilts the rotor forward at a certain angle. This is done by what is known as cyclic pitch change, i.e., changing the pitch of each blade once per revolution. More particularly, the angle of attack of each blade is increased every time it sweeps over the tail of the machine, thereby temporarily developing a greater amount of thrust than the other blades. The thrust developed by the rotor can be resolved into a vertical component (the actual lift that keeps the machine in the air) and a forward component (which propels the machine horizontally) (Fig. 3). Each blade can swivel about its longitudinal axis and its pitch is changed cyclically, through a linkage system, by a so-called swash-plate, which performs a sort of wobbling rotary motion around the shaft and swivels the blades to and fro as they rotate. The tilt of the swash plate can be varied by the pilot, and the tilt of the rotor follows the tilt of the plate.

The blade root hinges (called lag hinges) shown in Fig. 4 allow "blade flapping", as represented in Fig. 6. If there were no hinges, tilting of the plane of rotation of the rotor blades relative to the helicopter causes a periodic change in the speed of the blades. This would produce severe stresses in the blades; these stresses are relieved and cancelled by the hinge. Motion about the hinge enables the blade to rotate at constant speed irrespective of how much the rotor is tilted. In forward movement of the helicopter, the velocity due to blade rotation and that due to forward speed are added together, i.e., they intensify each other, on the advancing side of the rotor; on the retreating side, however, they are subtracted from each other. If the rotor blades were rigidly fixed to the shaft, the lift would vary cyclically and cause the helicopter to roll. This is prevented by the hinge. Instead, the blade flaps cyclically as it rotates.

The rotation of the rotor tends to cause the fuselage of the aircraft to rotate in the opposite direction (on the principle that any action calls forth a reaction). To prevent this, the single-rotor helicopter is provided at its tail with a small propeller producing a counteracting sideways thrust (Figs. 2, 3 and 7). Alternatively, the helicopter may have two rotors which revolve in opposite directions and thus counterbalance each other.

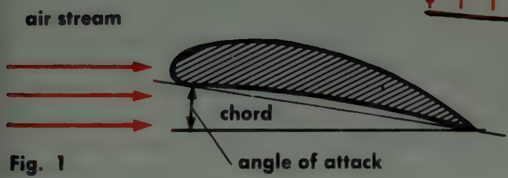


Fig. 1

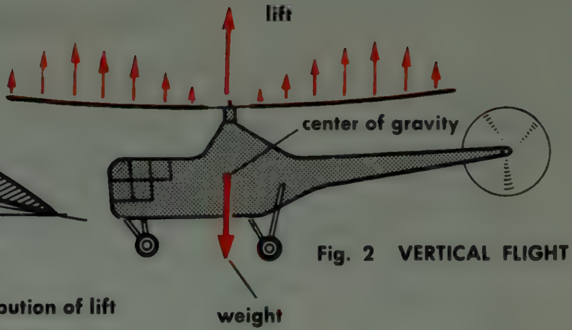


Fig. 2 VERTICAL FLIGHT

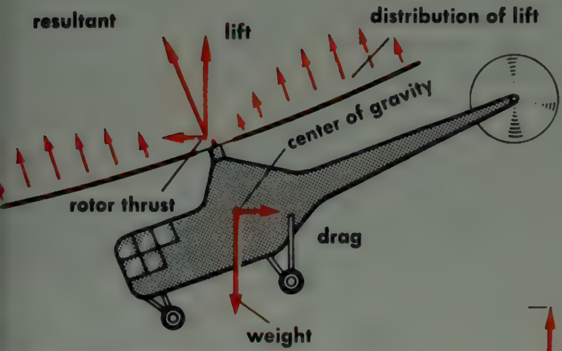


Fig. 3 FORWARD FLIGHT

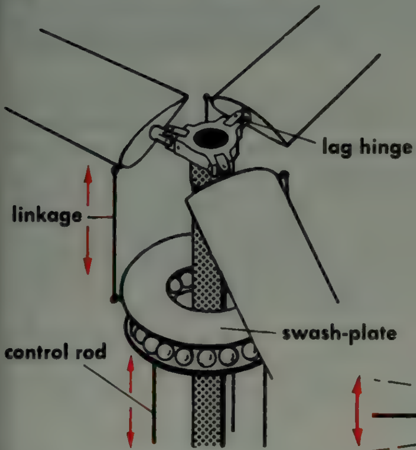


Fig. 4 ROTOR HEAD

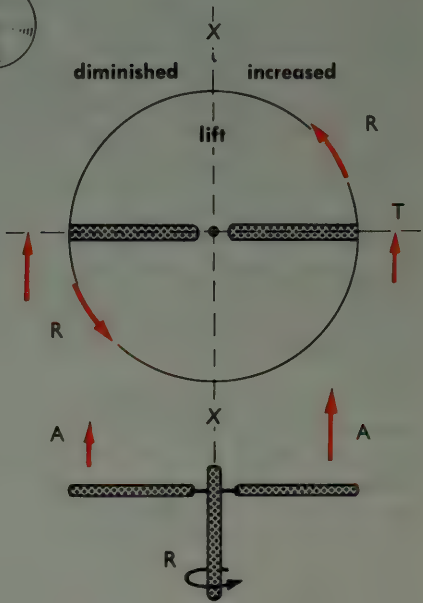


Fig. 5

Fig. 6

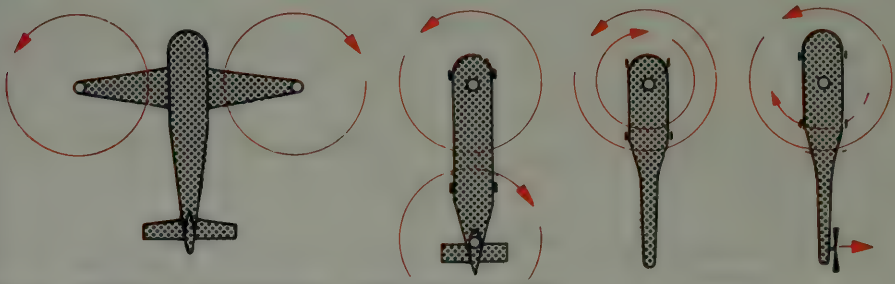


Fig. 7 METHODS OF PREVENTING FUSELAGE ROTATION

Jet engines for aircraft propulsion are of various types: turbojet, turboprop, ramjet, turbofan. The turboprop (turbo-propeller engine) is not strictly a jet propulsion engine. Rockets have also been used for the propulsion of aircraft and are jet propulsion devices, but they will be separately considered (see pages 312 and 314).

In jet engines of all types the air is taken from the outside, compressed, heated (by combustion of fuel), and then expanded in a jet or a turbine. The air is expelled from the jet at a much higher velocity than the intake velocity, and it is this increase in velocity that produces the desired propulsive thrust. It is based on the fundamental law of mechanics which states that action is equal to reaction. The compression energy developed in the combustion chamber of the jet engine is converted into impulse (= mass × velocity). The thrust results from the impulse of the air and combustion gases streaming out of the rear of the engine—not from any "pushing" against the air behind the aircraft. The law of action and reaction is illustrated by the familiar lawn sprinkler (Fig. 1): the water emerging from the jet at the end of the arm develops the "action"; the "reaction" force, of the same magnitude but in the opposite direction, causes the arm to rotate. In the case of a propeller-driven aircraft the propeller accelerates large quantities of air to a relatively low velocity, whereas in a jet-propelled aircraft the jet engine accelerates smaller quantities of air to far higher velocities (Fig. 2). Another example illustrating the principle of action and reaction is provided by a closed vessel (Fig. 3a) in which pressure is generated by the burning of fuel. If a hole is made in one side of this vessel, the combustion gases rush out at high velocity, and the vessel itself experiences a propulsive thrust in the opposite direction. The performance of a jet engine is proportional to the density of the intake air. It therefore diminishes with increasing altitude. However, as the drag on the aircraft likewise diminishes, there is actually an increase in speed at higher altitudes.

The simplest type of jet engine is the ramjet. It has no moving parts. Air entering the front of a tube shaped as shown in Fig. 4a is slowed down in the tube (because of the larger diameter of the middle part) and undergoes an increase in pressure in consequence. If fuel is injected into the centre of the tube and burned (Fig. 4b), the hot combustion gases flow out at high velocity from the rear of the tube (the nozzle of the jet). The velocity at the nozzle exit is higher than the flight speed of the aircraft; so a thrust is produced. The ramjet produces no thrust when the aircraft is stationary, because there is then no pressure build-up in the engine because the intake air has zero velocity. This type of jet is therefore efficient only at very high speeds.

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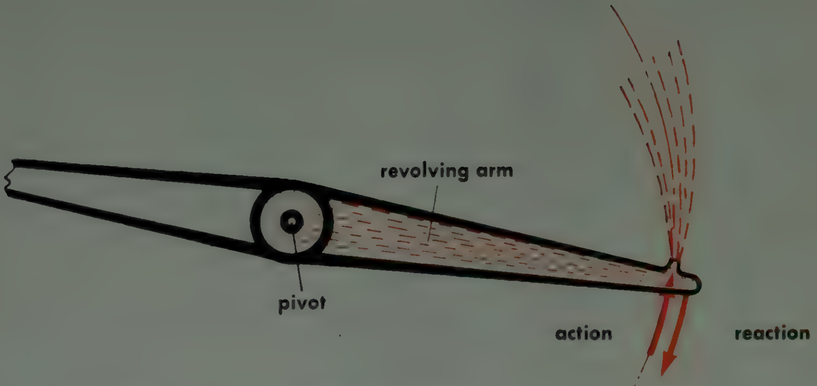
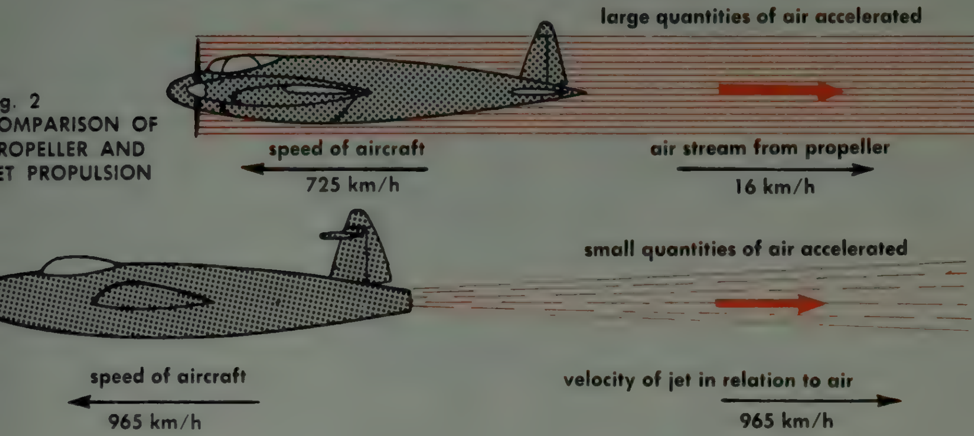
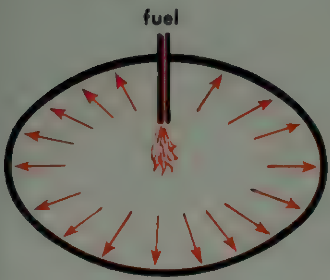


Fig. 1 PRINCIPLE OF ACTION AND REACTION



g. 2
OMPARISON OF
ROPELLER AND
ET PROPULSION



g. 3a HEATING: HIGHER PRESSURE

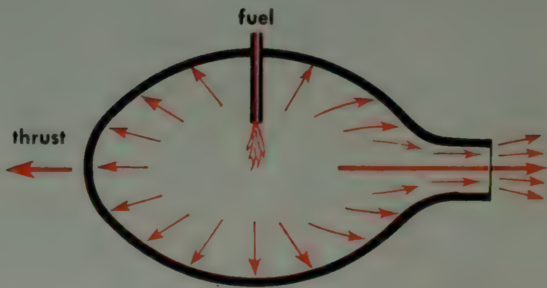


Fig. 3b ESCAPING GASES CAUSE THRUST

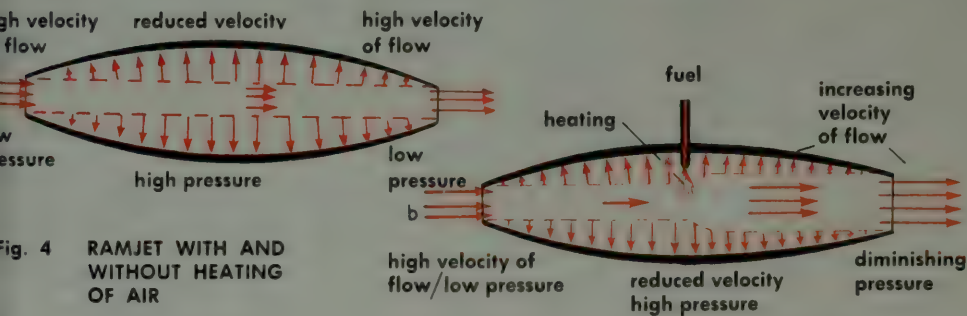
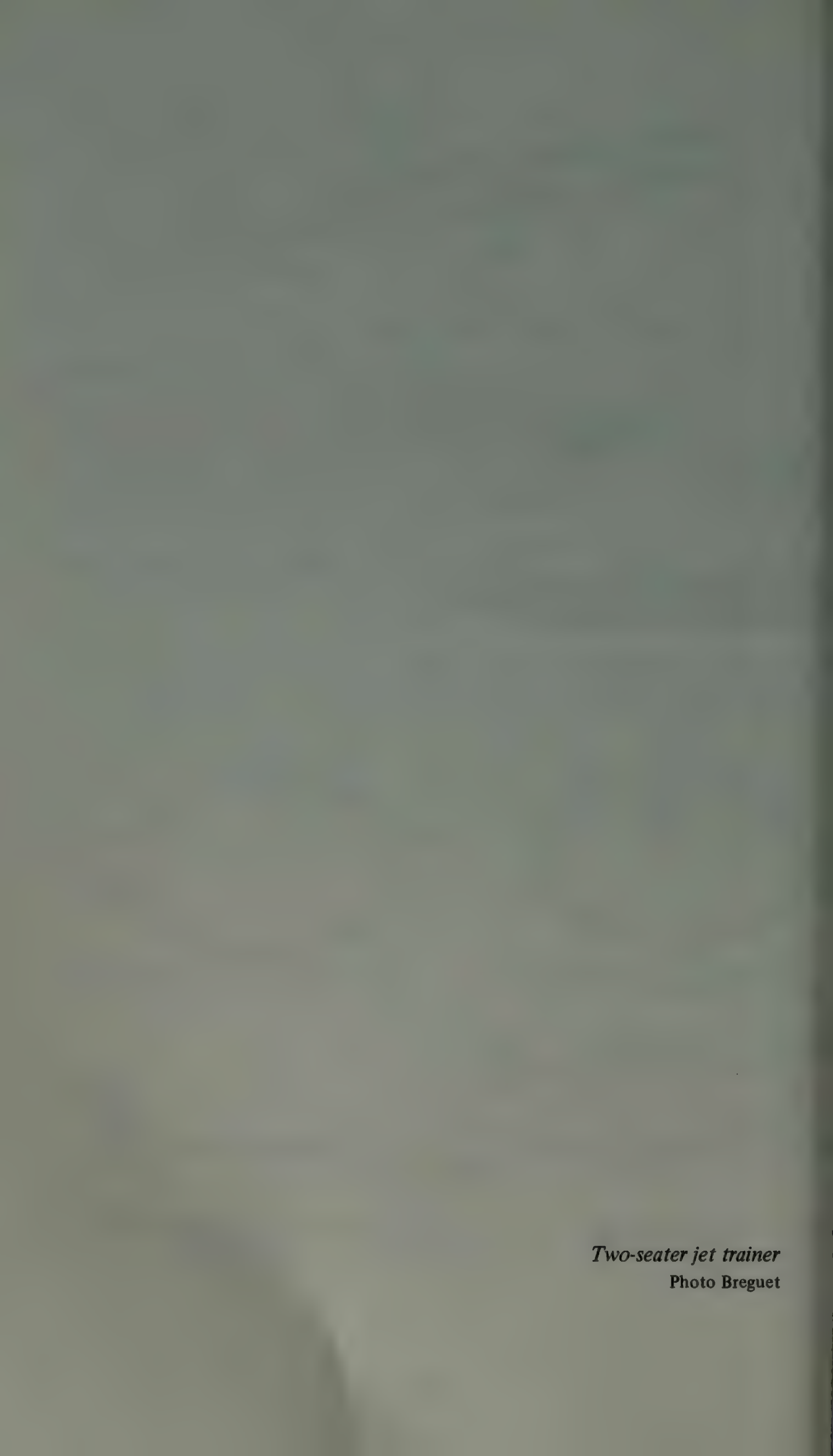
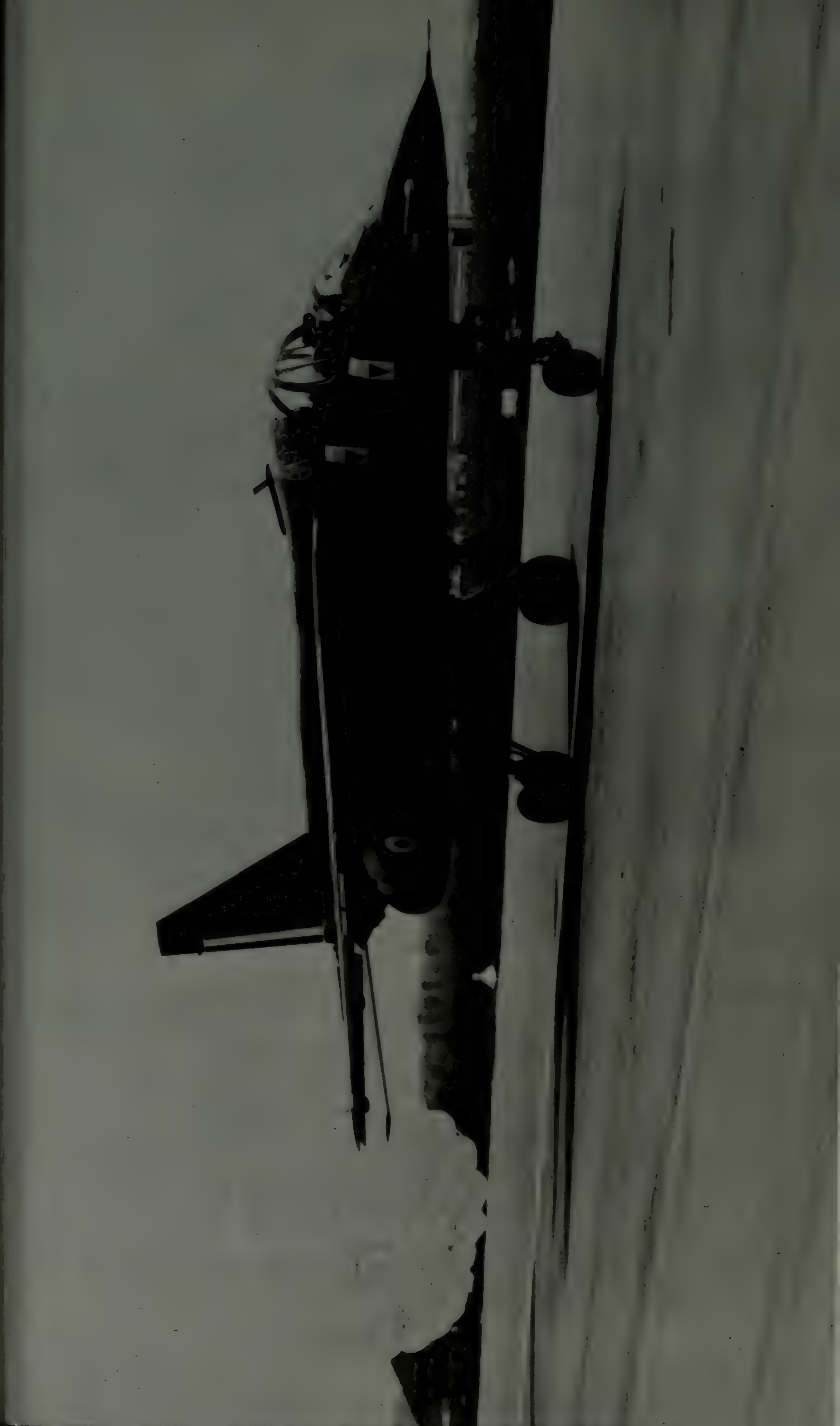


Fig. 4 RAMJET WITH AND WITHOUT HEATING OF AIR



Two-seater jet trainer
Photo Breguet



A turbojet (jet turbine engine) differs from the ramjet in that it is additionally provided with a compressor driven by a turbine (Fig. 1). A multi-stage axial compressor (comprising alternate sets of rotating and stationary blades) draws in the air and compresses it. Fuel is injected into the combustion chamber. The rise in temperature produces a considerable increase in the volume of the gases, which are expelled through the exhaust nozzle at the rear. The turbine absorbs only so much energy from the gases as is necessary to drive the compressor. The greater part of the energy is utilised to develop the propulsive thrust. The performance of a turbojet can be improved by means of an afterburner, which is a second combustion chamber placed between the turbine and the propulsion nozzle. Additional fuel is injected into this chamber and combustion is effected with the oxygen unconsumed in the main combustion process.

The propeller turbine engine (turboprop) likewise comprises a compressor, combustion chamber and turbine, but in this case nearly the whole of the energy is used to drive the turbine, which not only drives the compressor but whose excess power is used to drive a propeller (through reduction gears) (Figs. 2 and 3). The turboprop is much lighter than a reciprocating engine of equal power and is easier to construct in very large sizes (over 4000 h.p.). It is nevertheless a complex piece of machinery. The gases are very largely expanded in the turbine. The exhaust gases provide a certain amount of additional thrust.

A modified form of the turboprop has two concentric shafts revolving at different speeds. A turbine unit is mouned on each shaft. One shaft drives the compressor, the other drives the propeller. This type of engine is especially suitable for speed ranges which are too high for the conventional propeller-driven aircraft and too low for the turbojet. In the low-pressure compressor stage a proportion of the compressed air is bypassed and delivered to the propulsion nozzle (Fig. 4). Because of the higher power requirement of the turbine, the exit velocity of the gases is reduced, which is equivalent to increased propulsion efficiency. The low-pressure stage of the compressor is driven by the second turbine stage (shown black), and the high-pressure stage is driven by the first turbine stage (shown dotted).

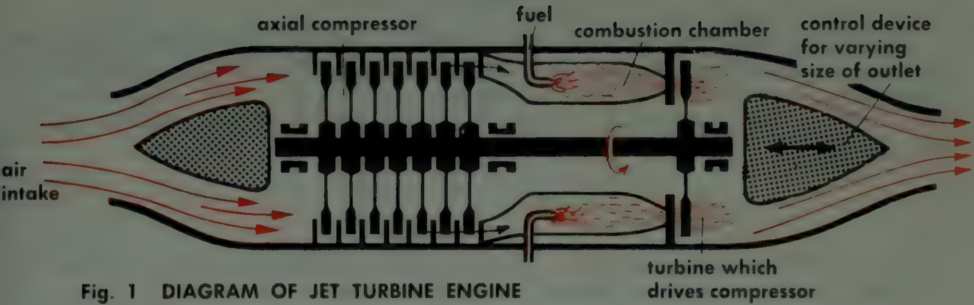


Fig. 1 DIAGRAM OF JET TURBINE ENGINE (*turbojet*)

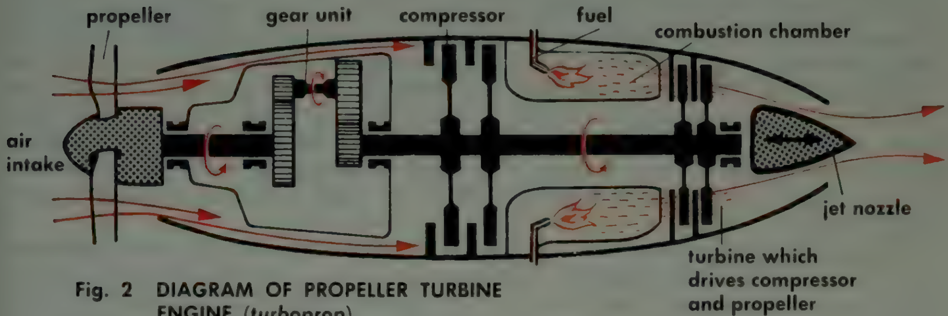


Fig. 2 DIAGRAM OF PROPELLER TURBINE ENGINE (*turboprop*)

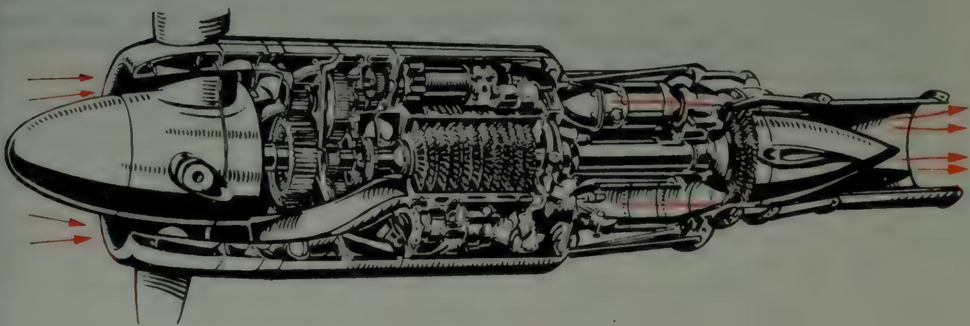


Fig. 3 TURBOPROP ENGINE

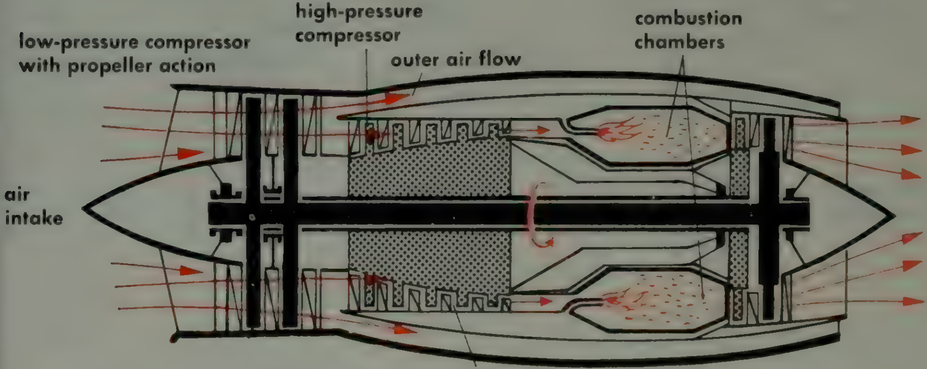


Fig. 4 DIAGRAM OF TWO-STAGE TURBOPROP

When a parachute has opened and is falling steadily, a state of equilibrium exists in which the resultant force of drag and lift acting upon the falling parachute is equal to the weight (Fig. 2). The magnitude of the drag force (W) is dependent on the square of the velocity of descent (v), the magnitude (F) of the projected area of the parachute, and the specific gravity (y) of the air:

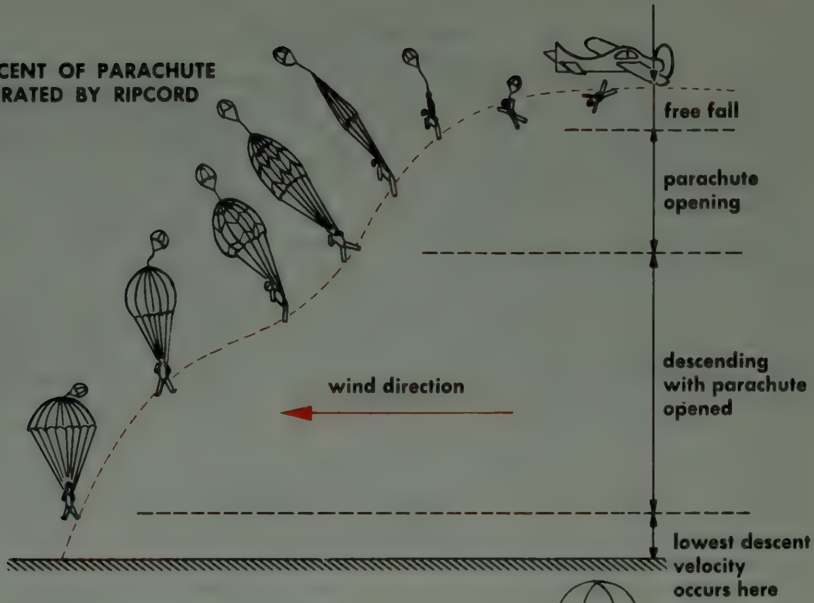
$$W = \frac{\gamma}{2g} . v^2 . F.$$

If the angle of the approach velocity becomes zero, the drag force will be equal to the weight, while the rate of descent remains the same.

A parachute descent is conceived as comprising various zones (Fig. 1). In the free fall zone the parachutist falls with ever increasing velocity until a certain limiting velocity—determined by air drag and friction—is reached. The ripcord is then pulled causing the pilot parachute to open; this is a small auxiliary parachute attached to the top of the main parachute; it pulls the latter out of its pack. When the main parachute opens, the velocity of fall is reduced from about 110 m.p.h. to about 10–15 m.p.h. This sudden slowing-down produces the "opening shock". The further descent is then made at constant velocity. This velocity depends not only on the diameter and shape of the parachute, but of course also on the parachutist's weight, the air permeability of the parachute fabric, and the density of the air. For a normal rate of descent the impact on landing is about the same as that on jumping from a height of 5 ft.

The parachute may be opened by means of a ripcord pulled by the parachutist or by an automatic timing device. The ripcord withdraws the rip pin which secures the flaps on the parachute pack (Fig. 3). The ripcord system is used for pilot escape. For mass jumps by parachute troops, a static line attached to the aircraft pulls the parachute out of its pack. The principal parts of a parachute are the canopy, suspension lines, chute pack, pilot parachute, and harness (Fig. 4). Modern parachutes are made of nylon. A typical pilot-escape parachute has a diameter of 28 ft. Airborne troops generally use parachutes of somewhat larger diameter. The harness is so designed that it can instantly be released from the wearer's body.

Fig. 1 DESCENT OF PARACHUTE OPERATED BY RIPCORD



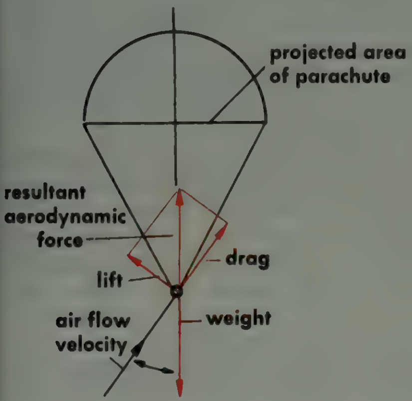


Fig. 2 FORCES ACTING UPON A PARACHUTE

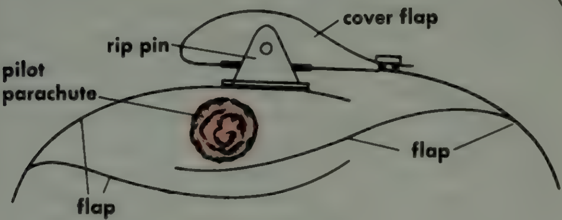


Fig. 3 HOW PILOT PARACHUTE IS PACKED

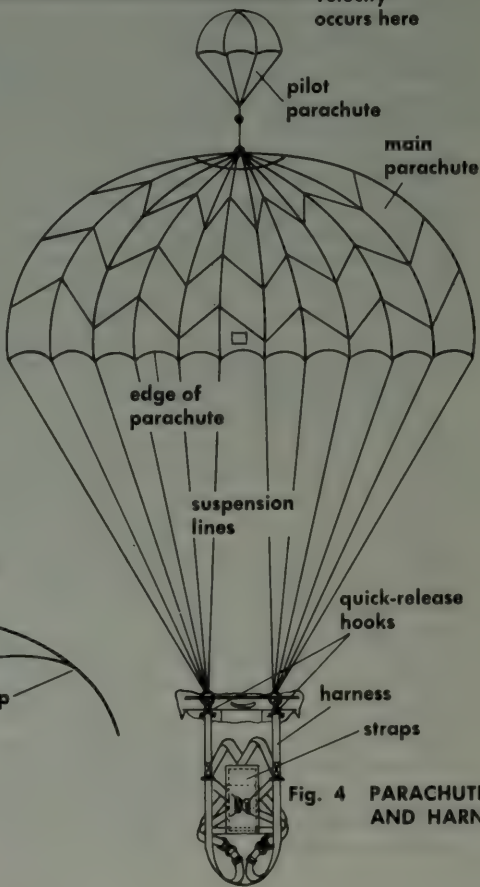


Fig. 4 PARACHUTE AND HARNESS

The *altitude* of an aircraft above sea level is measured by means of the altimeter, which is essentially an aneroid barometer (p.252, vol.I) and provides an indication of atmospheric pressure which is calibrated—by means of a standard formula—to give readings direct in terms of altitude (Fig. 1). The instrument is affected by variations of barometric pressure, and provision is made for adjustment. The temperature also affects the altitude readings obtained. The correct use of the terms "altitude", "height" and "elevation" is indicated in Fig. 2. By setting a pointer on a subsidiary scale of air pressure values the altimeter can be made to indicate the actual elevation of the airport when the aircraft has landed on the runway. Also, by means of this adjustment the altimeter can be made to give actual readings of height above airport runway level, i.e., the "datum" (reference level) in relation to which the height is measured can be varied. This vertical adjustment of the relative height indications is represented by the full and the dotted red lines in Fig. 3. The setting of the instrument in any specific case will also depend on the atmospheric pressure existing at that particular time and place. The change-over from altitude readings to height readings is effected prior to landing. This is done on entering (from above) the statutory zone (generally at least 1000 ft. in depth) shown shaded by oblique lines in Fig. 3. When the aircraft takes off, the opposite adjustment is made on entering this zone (from below).

The height of an aircraft above ground level is additionally measured by means of a *radio altimeter.* A radio signal is transmitted to the ground, reflected back, and picked up by a receiving instrument (Fig. 4). The length of time between transmission and reception of the signal is evaluated by electronic means and provides an indication of the height. The readings obtained in this way are reliable only over flat country and over water. Surface irregularities cause inaccuracies.

The *rate-of-climb indicator* (variometer, vertical speed indicator) is a pressure gauge whose dial indicates the rate of climb or descent in thousands of feet per minute. It comprises a differential pressure element (usually a diaphragm capsule) one side of which is connected to the ambient static pressure and the other side is connected to this pressure through a constriction (a narrow orifice or the like). A change in altitude of the aircraft is associated with a change in static pressure. Because of the constriction, equalisation of pressure can only take place slowly. The differential pressure at the construction provides an indication of the vertical speed.

The direction of the flight path of an aircraft is measured in terms of the angle (the course) between the tangent to the path and a particular reference direction, e.g., compass north, true north or grid north (Fig. 6). The angle between this reference direction and the longitudinal axis of the aircraft is called the heading. If there is no wind, the two angles, i.e., the course and the heading, will be identical. To compensate for winds blowing obliquely in relation to the flight path, the heading will differ from the course (the latter being the actual desired direction). The shortest flight path between two points A and B is not the loxodrome (this being the path obtained if the course is kept constant throughout the flight: cf. page 278) but a great circle (a circle on a sphere—more particularly: the earth—whose plane passes through the centre of the sphere). In practice, the navigator determines the great circle passing through A and B and then approximates it by a "polygon" of stretches of loxodrome which can be flown with the aid of a magnetic compass or a gyrosyn compass. The gyrosyn compass system comprises a directional gyroscope (Fig. 7), a compass indicator, and a so-called flux valve (a detector of magnetic field). The directional gyroscope ensures directional stability of the compass system; the flux valve senses the earth's magnetic field and transmits signals which correct any wandering of the gyroscope.

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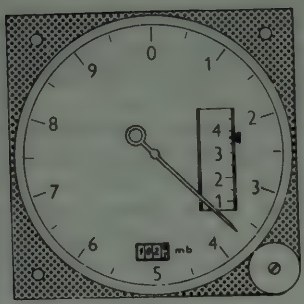

Fig. 1 ALTIMETER

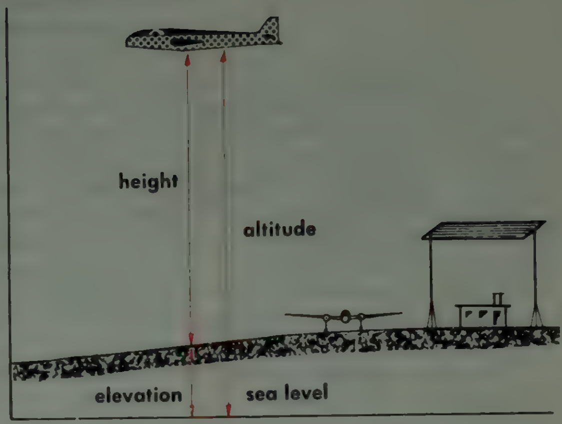

Fig. 2

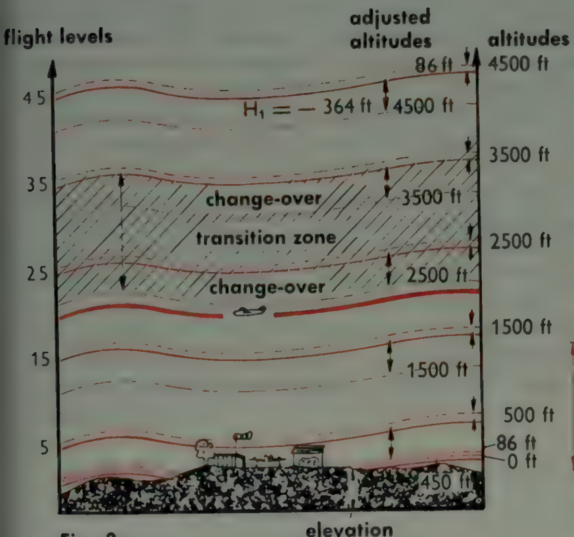

Fig. 3

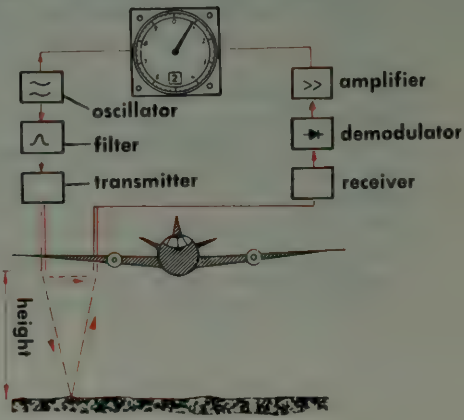

Fig. 4 RADIO ALTIMETER

Fig. 5 RATE-OF-CLIMB INDICATOR

Astronomical aircraft navigation is done with the aid of the astro-compass, which is a non-magnetic instrument which enables the true heading of an aircraft to be found by sighting upon a celestial body. The readings obtained with this instrument can be used for correcting the directional gyroscope from time to time.

With reference to the "speed" of an aircraft in flight a distinction is made between the airspeed (the speed relative to the air through which the aircraft moves), the groundspeed (the speed relative to the ground), and the wind velocity or speed. These speeds are represented by v_e, v_g and v_w respectively. If two of these are known, the third speed can be found as a resultant in the so-called wind triangle (Fig. 8) in which the length and direction of each vector ("arrow") corresponds to the magnitude and direction of the speed concerned.

The airspeed is generally measured by means of the *airspeed indicator.* It operates by utilising the pressure difference between static and dynamic air pressures, the instrument usually being calibrated to give readings in knots or in miles-per-hour. The dynamic pressure is produced in a so-called pitot tube (whose open end points in the direction of flight), the pressure difference being measured by means of a diaphragm capsule (Fig. 9). The instrument shows true airspeed under standard sea-level conditions (atmospheric pressure 29.92 inches of mercury, temperature 15° C). In actual practice, corrections must be applied for varying atmospheric conditions. For navigational purposes it is important to know the groundspeed, and this can be determined by means of the wind triangle if the local wind velocities are known. There are also instruments for the direct measurement of groundspeed.

In so-called contact flight the pilot establishes his position with the aid of prominent landmarks (bridges, railroads, mountains, etc.) or—at night—air beacons and other lights. However, in modern high-level flying this method has been largely superseded, the establishment of position being achieved by means of navigational instruments. For night flying over uninhabited areas, the periscope sextant (Fig. 10) is used for observing the altitudes of celestial bodies. To make an observation, the periscope tube of the instrument is extended so as to protrude out of the fuselage of the aircraft.

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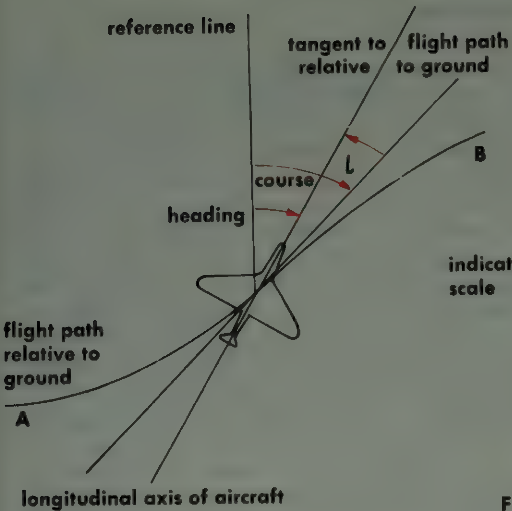

Fig. 6

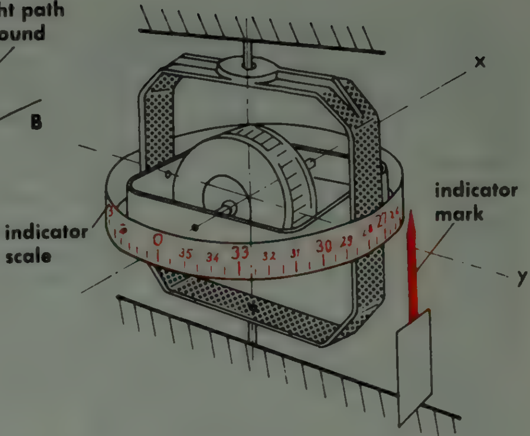

Fig. 7 DIRECTIONAL GYROSCOPE

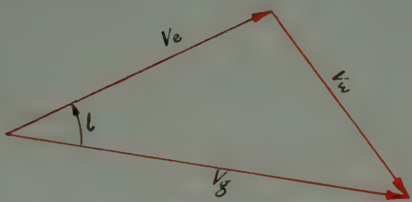

Fig. 8 WIND TRIANGLE

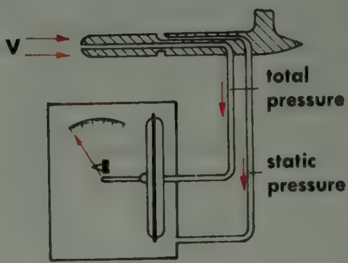

Fig. 9 PITOT TUBE

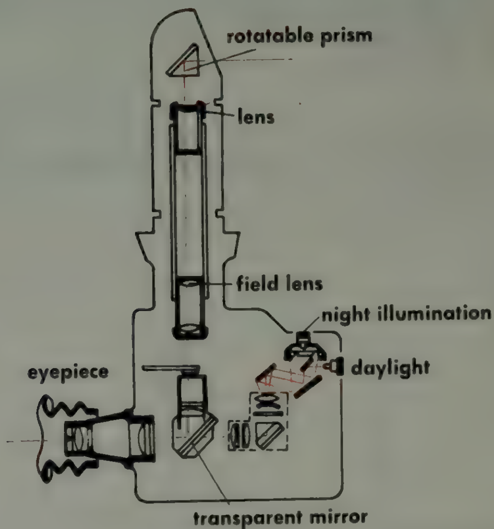

Fig. 10 PERISCOPE SEXTANT

The most widely used means of establishing an aircraft's position is by the use of *electronic aids.* Radio direction-finding (DF) equipment relies on the direction property of a loop antenna. By rotating the antenna it is possible to establish the direction from which the signals from a ground-based radio beacon are coming and thus to find the aircraft's bearing in relation to that direction, i.e., the angle between the direction of the signals and the longitudinal axis of the aircraft. Since the aircraft's heading (see above) is also known, it is possible to establish the angle between the radio signal direction and, for example, true north. The principle is illustrated in Fig. 11. By taking bearings on three different radio beacons it is possible to establish the position of the aircraft. In a modern aircraft the equipment usually takes the form of an *automatic direction finder* (radio compass) whose antenna is kept directed at a particular radio beacon by means of an electrical sensing circuit. The antenna position, and therefore the aircraft's bearing in relation to the signal direction, is indicated on a dial (Fig. 12). In the navigational system known as VOR (very high-frequency omnidirectional range), which is used chiefly in the United States, a direct indication of bearing from the transmitting station is obtained. Other well known navigation systems are Decca and loran (long range navigation), developed in Britain and in the United States respectively.

Inertial guidance systems are entirely independent of outside signals. For this reason they are valuable for missible control and submarine navigation. The operation is based on the measurement of the amount and direction of acceleration, usually in three mutually perpendicular directions. The speed at any particular instant and the distance travelled are obtained by a process of integration which is carried out with the aid of computer equipment (Fig. 13). Fixed reference axes are established by three gyroscopes. These are mounted on gimbals with mutually perpendicular axes. Deviations from the reference co-ordinates are fed to a computer.

The *instrument landing system* (ILS) is illustrated in Fig. 15. Radio signals are transmitted from two ground-based antenna systems and provide horizontal and vertical guide planes respectively. The pilot's indicating instrument (Fig. 14) comprises a pointer for horizontal and another for vertical guidance. Approximate indications of distance are obtained on passing the outer marker (OM), middle marker (MM) and boundary marker (BM) transmitters. The signals are converted to acoustic and visual indications. In another system the aircraft is tracked by ground radar and the pilot receives instructions by radio (he is "talked down"). This is known as ground controlled approach (GCA).

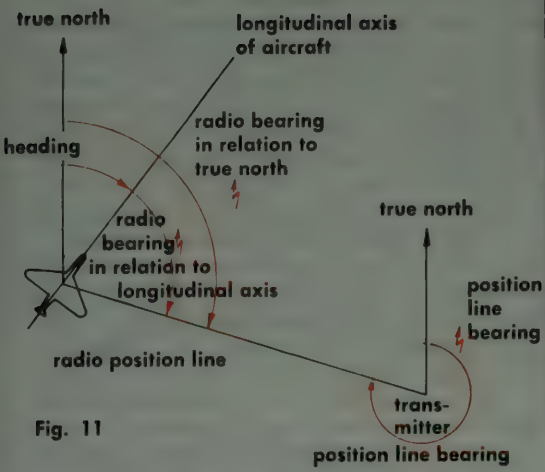

Fig. 11

Fig. 12 RADIO COMPASS AND VOR

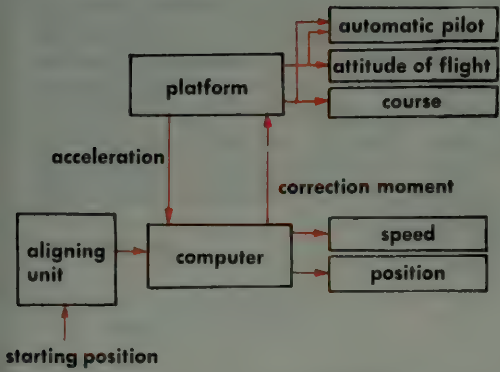

Fig. 13 BLOCK DIAGRAM OF INERTIAL GUIDANCE SYSTEM

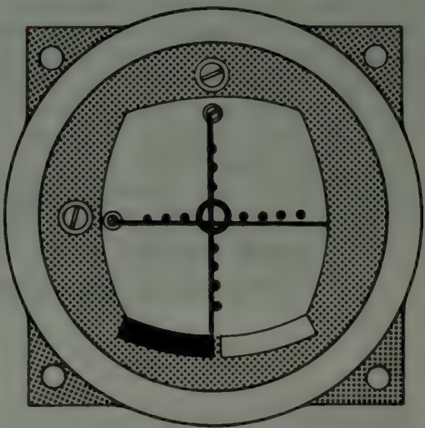

Fig. 14 INDICATING INSTRUMENT OF INSTRUMENT LANDING SYSTEM (ILS)

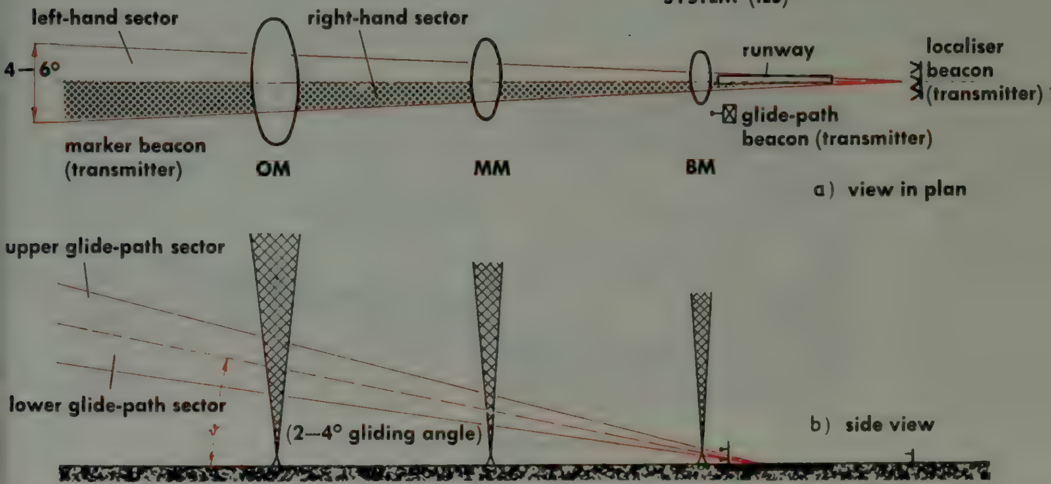

Fig. 15 INSTRUMENT LANDING SYSTEM

Hovercraft belong to the category of what can be described as "air cushion vehicles" or "ground effect vehicles". In principle, a hovercraft comprises a body or hull in which a rotor (lift fan) is so mounted that it can produce an air cushion on which the craft is supported. There is therefore no contact between the craft and the ground and no friction to overcome. It can travel on land as well as on water. The hovercraft relies on the so-called ground effect to form the air cushion. The principle underlying this effect is illustrated in Fig. 1. The flow of air produced by a fan close to the surface of the ground acquires an annular pattern: all the flow and turbulence is concentrated at the edges, while at the centre an air cushion is formed in which the air is very nearly at rest. At the edge of this cushion an annular jet of air develops which forms a curtain that insulates the cushion from the surrounding lower-pressure atmospheric air.

A number of devices based on the hovercraft principle have been constructed and tested. Fig. 2 shows a very simple machinery comprising a lift fan and a large chamber (plenum chamber) in which the air cushion is formed. The "hovering" distance above the surface of the ground in this case is dependent on the relation between the air intake and air discharge rate. However, as the pressure in the chamber is low, it is necessary to provide a large base area in order to obtain a reasonable amount of lift.

A more efficient arrangement in this respect is the annular chamber (Fig. 3), which directly produces the annular "jet" envisaged in Fig. 1. The effect is improved by sloping the peripheral nozzles inwards. As a rule, one or more horizontal-thrusting propellers are fitted for forward propulsion of the hovercraft. Extensive tests have shown that these craft give the best performance and stability when the "hovering" height is somewhat less than one-tenth of the diameter, i.e., a craft of 20 ft. diameter should operate at a height of about 2 ft. above the surface of the ground. To reduce wind resistance, however, hovercraft are usually not constructed circular on plan—though this would be the most favourable shape from the stability point of view—but are made oval or rectangular. For example, Fig. 4 illustrates a projected large hovercraft which will be provided with an annular air jet. It will have a total weight of 400 tons and be capable of speeds of up to 100 m.p.h.

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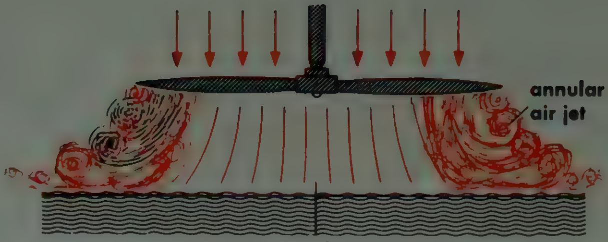


Fig. 1 GROUND EFFECT (*formation of air cushion and annular jet*)

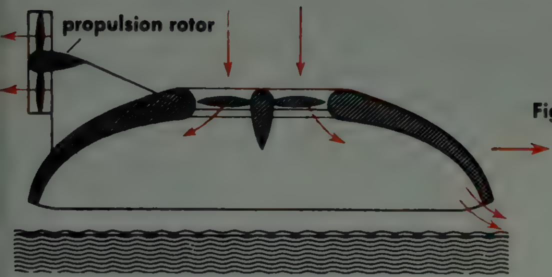


Fig. 2 HOVERCRAFT WITH PLENUM CHAMBER (*propulsion by jet impulse*)

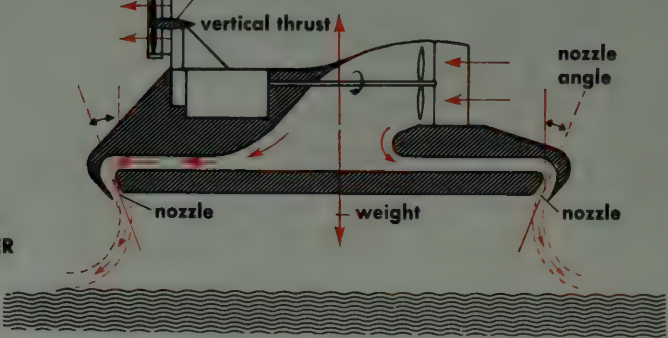


Fig. 3 HOVERCRAFT WITH ANNULAR CHAMBER

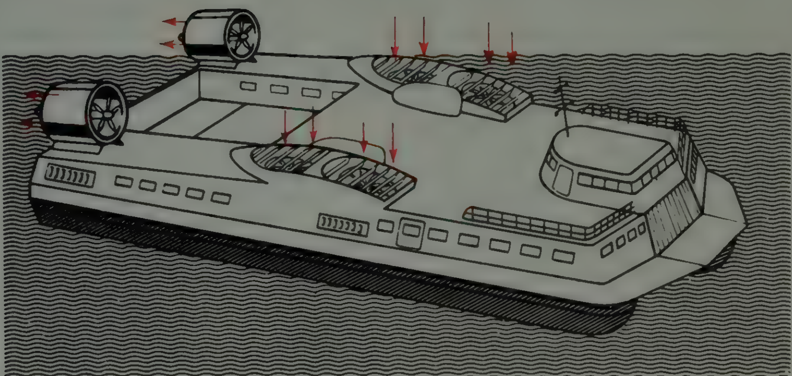
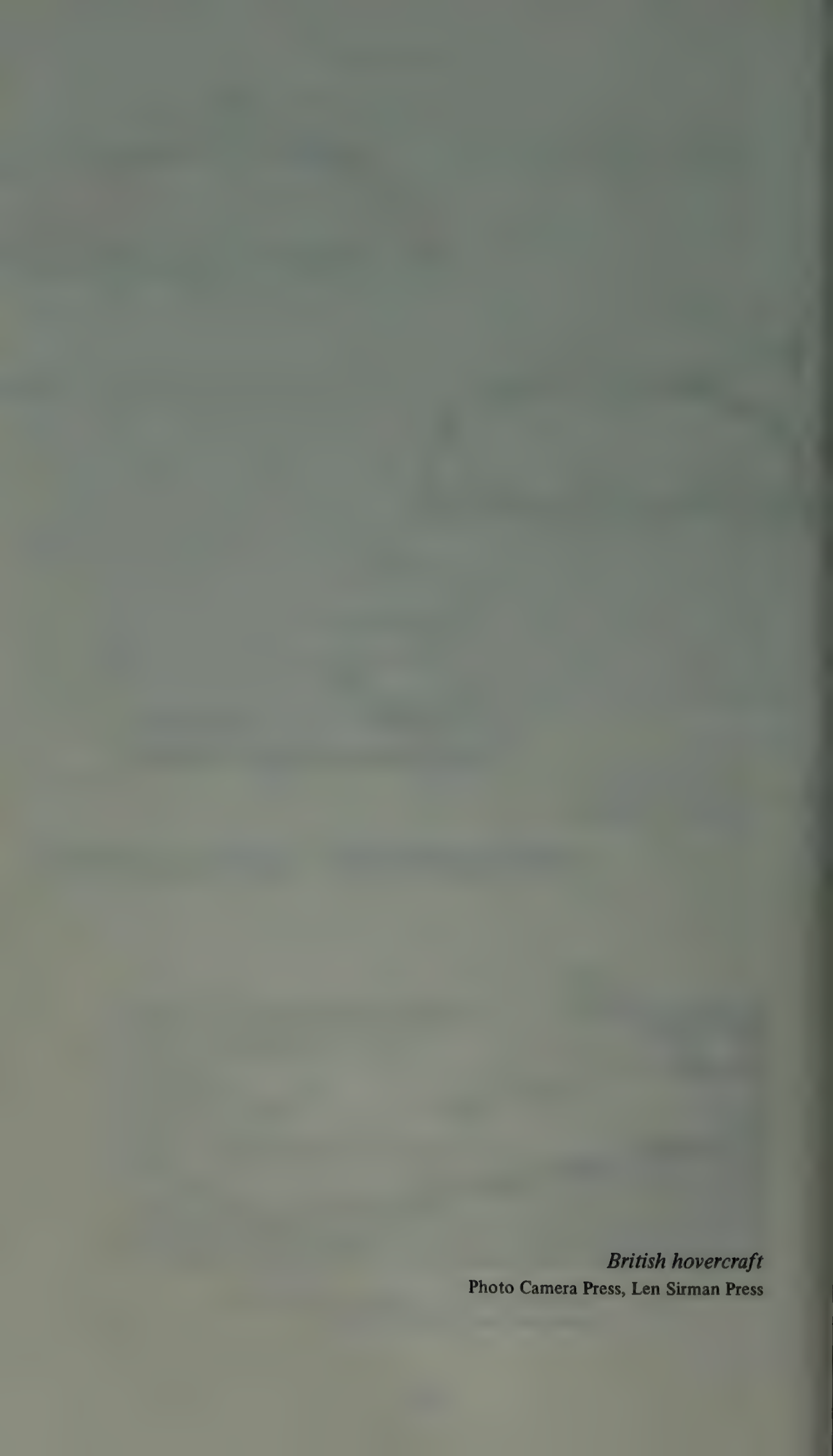


Fig. 4 PROJECTED LARGE PASSENGER HOVERCRAFT (*100 m.p.h., total weight 400 tons*)



British hovercraft
Photo Camera Press, Len Sirman Press

The forces acting upon a hovercraft are indicated in Fig. 5. With increasing forward speed the aerodynamic drag plays an increasingly important part. In addition to this pure form drag, which is also associated with the forward motion of aircraft, there is an impulse drag which is caused by the acceleration of the air needed for forming the supporting cushion under the hovercraft. At low speeds this impulse drag is greater than the form drag, at high speeds the latter is greater. For forward motion the rear nozzles on the underside of the craft can be so adjusted as to develop a rearward thrust equivalent to about 20% of the total thrust necessary for propelling the craft, the other 80% being supplied by the horizontally acting propeller.

A more elaborate system of sealing the air cushion is employed in the craft illustrated in Figs. 6 and 7. Here the insulating "curtain" is formed by a circulating stream of air at the edge. Higher efficiency and reduced air losses from the cushion are claimed for this system. On the other hand, the cost of construction is higher.

Braking and steering a hovercraft are performed aerodynamically. Simple control surfaces (rudders) such as those used on aircraft may be useful at high speeds, but are ineffective at low speeds. Hence these craft are steered by thrust impulses, e.g., by means of lateral propellers. The same principle is employed for braking: the forward propulsion propellers are of the variable-pitch type; by appropriately varying the pitch angle of the blades to a negative value, these can be made to develop a backward thrust. In the design of hovercraft the problem of stability plays a major part. For example, with increasing speed of the craft its stability is impaired by the increasing dynamic pressure which tends to break down the supporting air cushion. The dynamic pressure must on no account exceed the pressure existing within the air cushion.

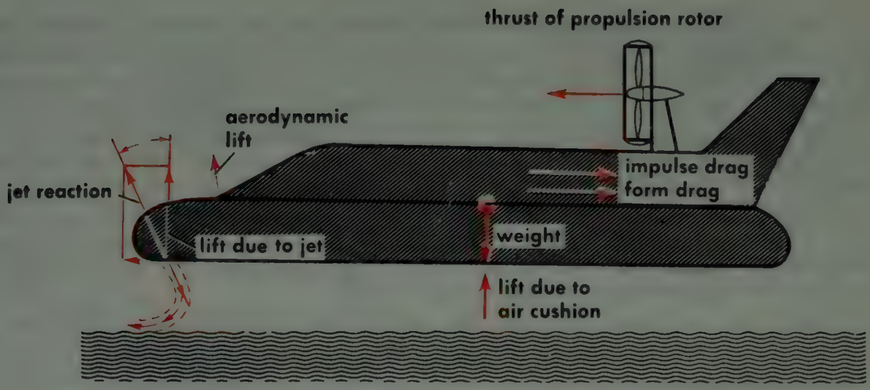


Fig. 5 FORCES ACTING ON A HOVERCRAFT IN MOTION

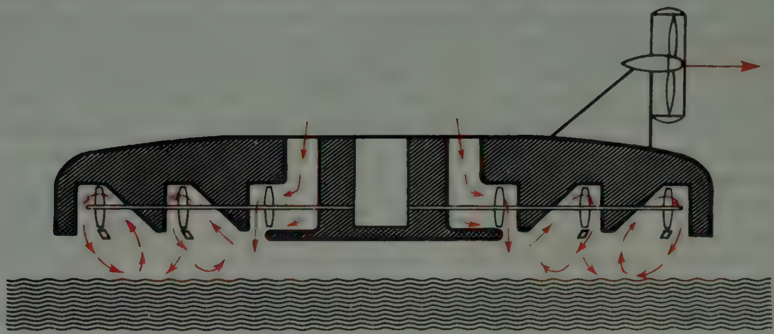


Fig. 6 HOVERCRAFT WITH AIR CURTAIN
(*Weiland two-stage system*)

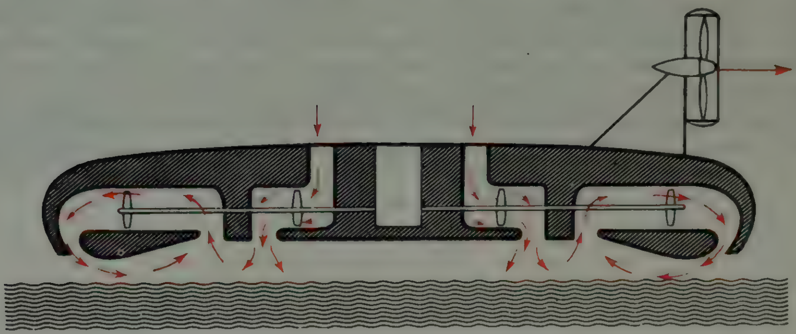


Fig. 7 HOVERCRAFT WITH AIR CURTAIN
(*Miller single-stage system*)

A rocket is a jet-propelled missile which carries the source of its propulsive energy along with it and whose functioning is independent of the presence of an atmosphere. It is more particularly in this latter respect that a rocket differs from an aircraft jet engine (see page 292 f.). Whereas an "air-breathing" jet engine uses air to burn its fuel and can therefore operate only in the atmosphere, the propulsion of a rocket requires no external aid to burn its fuel. For this reason the altitudes and speeds attainable by rockets, as distinct from those attainable by jet aircraft, are in effect "unlimited". Every jet propulsion system is based on Newton's fundamental law of "action" and "reaction", which states that every "action" produces a "reaction" of the same magnitude but acting in the opposite direction. Some elementary examples of this principle are illustrated in Fig. 1.

When the air is allowed to escape from an inflated balloon in one direction (action), the balloon will move in the opposite direction (reaction) (Fig. 1a). Something similar happens when a gun is fired (Fig. 1b). In order to accelerate the projectile to the required muzzle velocity, the force due to gas pressure developed by the explosive must act for a certain time. This force multiplied by the length of time is called the "impulse"; it is equal to the mass of the projectile multiplied by its muzzle velocity.

Similar conditions also apply to rocket propulsion (Figs. 1c and 1d). When a hot gas under high pressure is produced by the burning of a rocket fuel (propellant) in a combustion chamber, the gas will exercise an equal pressure in all directions upon the walls of the chamber. Now if an opening is made in one side of the chamber, the gas will come streaming out at supersonic velocity; at the same time a reaction force will be exerted on the opposite side of the chamber, and it is this reaction force that thrusts the rocket forward. This propulsive force (thrust) is equal to the mass of propellant discharged per unit of time multiplied by the flow velocity. This means that the thrust becomes greater in proportion as the gas discharge rate per second and the flow velocity are higher. The thrust is therefore not produced—as is often erroneously supposed—by the outflowing gases "pushing" against the surrounding medium (atmosphere), but is developed purely as a reaction force due to the expulsion of matter at high velocity from a closed system. Rocket propulsion is therefore the only propulsive system that can function in vacuum. Fig. 2 shows some practical applications of the impulse theorem and the principle of reaction.

When the entire propellant supply carried along by the rocket has been transformed into thrust, the rocket has attained its maximum velocity. This final velocity depends upon the flow velocity of the gas and the so-called mass ratio of the rocket, i.e., the ratio of the initial mass (mass of the rocket at blast-off, including the propellant supply) to the final mass (the initial mass minus the mass of the consumed propellant). An increase in final velocity is attainable by means of the multi-stage rocket. Each stage has its own propulsion system and propellant supply; the final stage additionally carries the payload. When the propellant supply of one stage has been exhausted, this stage is mechanically detached from the subsequent stages. The thrust developed by the next stage serves to produce a further increase in the velocity of the rocket, now reduced in mass after having discarded the previous (spent) stage.

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Fig. 1a GAS-FILLED BALLOON

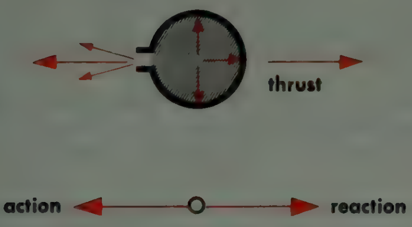


Fig. 1b GUN

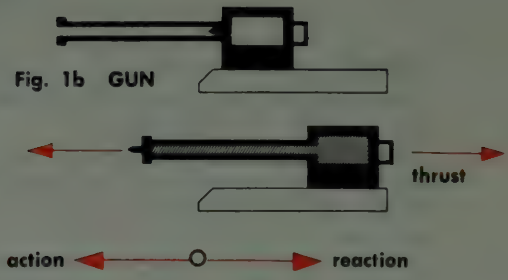


Fig. 1c COMBUSTION CHAMBER

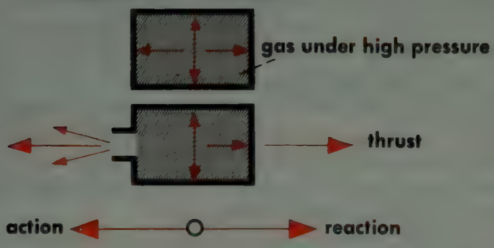
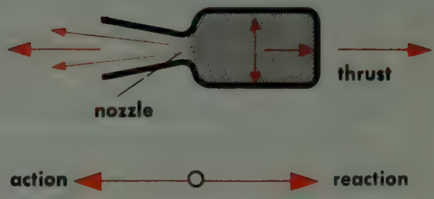


Fig. 1d ROCKET



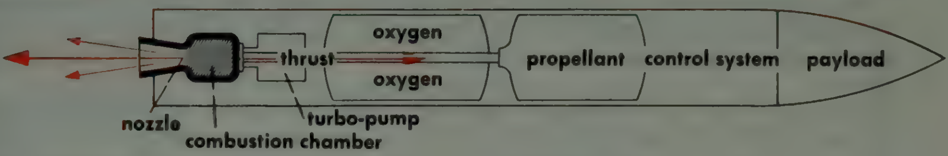
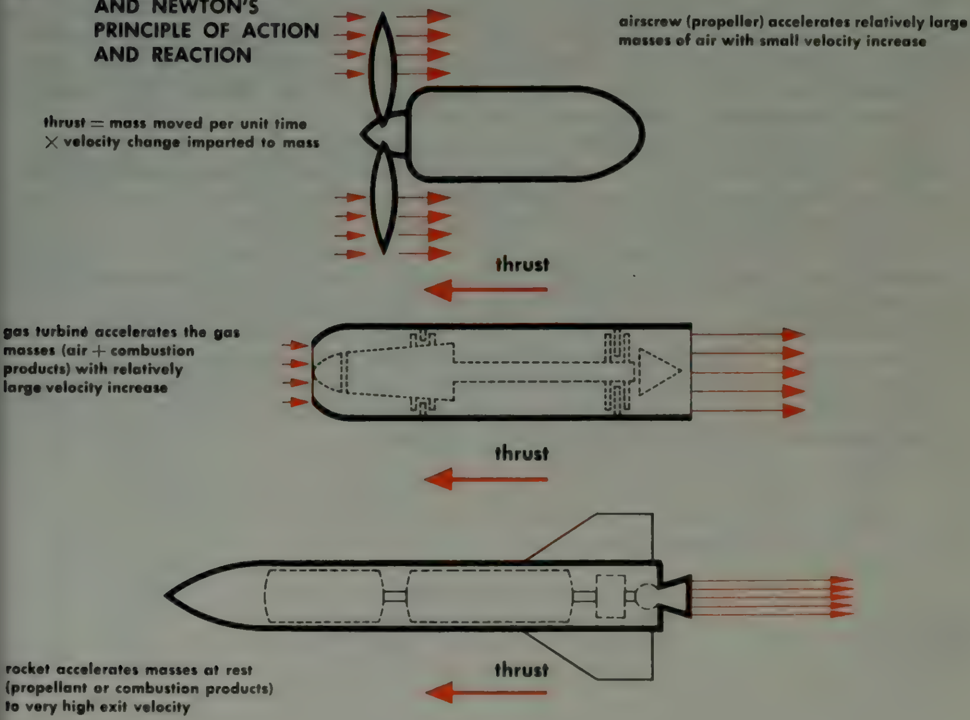


Fig. 2 IMPULSE THEOREM AND NEWTON'S PRINCIPLE OF ACTION AND REACTION



Rocket propulsion systems:

Acceleration of the mass of a propellant to high flow velocities in a rocket motor can be achieved in various ways.

In chemical-propellant rockets, as at present employed, hot gases under high pressure are produced in a combustion chamber and acquire their velocity in a nozzle. Rockets of this class can be subdivided into solid-propellant, liquid-propellant, and hybrid rockets.

In the case of a solid-propellant rocket the propellant, which consists of the combustibles and an agent supplying the oxygen for its combustion and which may have a variety of surface configurations and arrangements, is introduced into the combustion chamber, where it burns. In so doing, it produces a hot high-pressure gas which is discharged through a nozzle and thus produces the thrust that propels the rocket. In liquid-propellant rockets the liquid combustibles are contained in tanks and fed into the combustion chamber through an injector head by a propellant supply system. Most liquid-propellant rockets use two combustibles (bipropellant system) such as liquid oxygen and kerosene. Also, a liquid oxygen supplying agent can be used in conjunction with a solid combustible (hybrid rockets), whereby better control of the thrust developed by the rocket motor is obtained.

In another method (still in the experimental stage) the hot gas is produced, not by chemical combustion, but with the aid of nuclear energy. This method involves heating a working fluid (e.g., water or hydrogen) in a nuclear reactor (p.68, vol. I), the gas being discharged through a nozzle in which it is accelerated and develops the thrust.

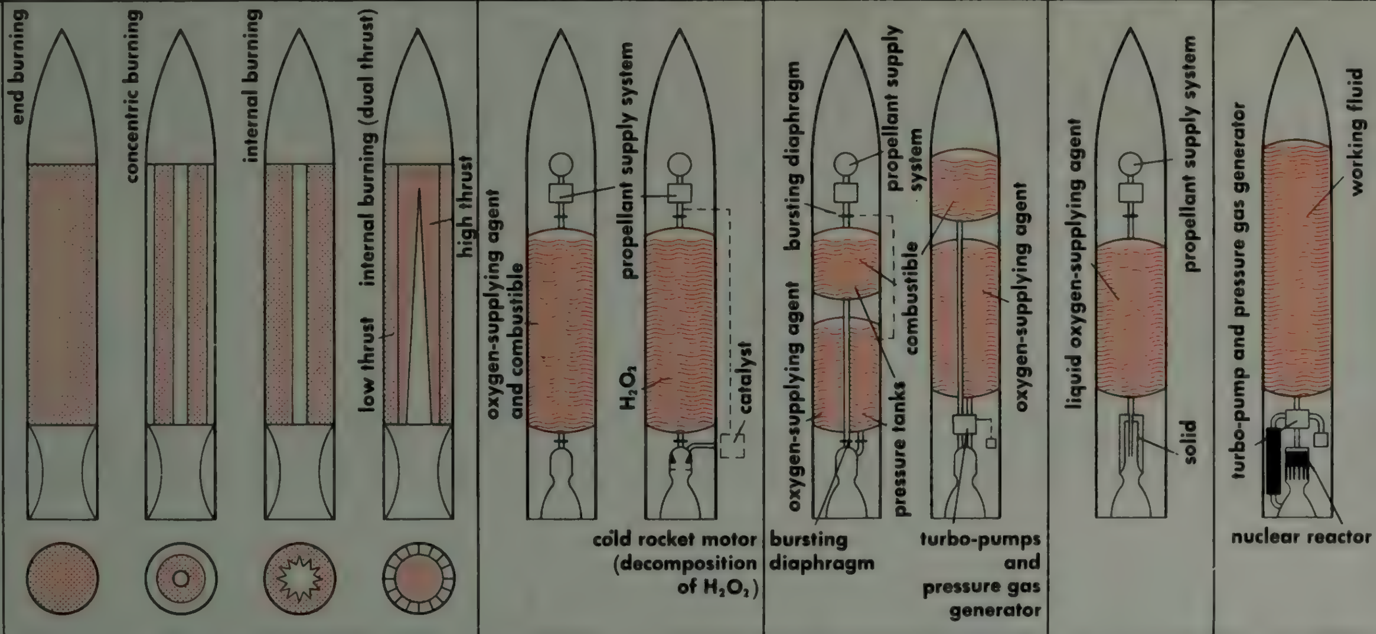
Whereas chemical-propellant rockets are characterised by high thrust for short durations and are therefore more particularly suitable for getting large payloads launched from the ground, electrical propulsion systems have been proposed which would yield a relatively low thrust over a long period of time. Such systems would be suitable for space travel, i.e., under interplanetary conditions of rocket flight.

Of the various electrical propulsion systems that have been proposed, the one most closely resembling the chemical system is the electro-thermal arc jet. This device utilises an electric arc to heat a working fluid (e.g., hydrogen, helium, etc.), which is "thermodynamically" accelerated in a nozzle.

In another system, the so-called ion rocket, electrostatic fields are used to accelerate positively-charged ions to very high specific impulse values. The system comprises an ion-producing and an ion-accelerating device, as well as a neutralising zone where electrons are added in order to restore the electrical equilibrium.

In yet another electrical propulsion system the plasma produced by an electric arc is accelerated by means of an electromagnetic field. ("Plasma" is the term applied to a gas with more than 50% of its particles ionised, i.e., electrically charged). Also, there is a somewhat different type of thrust device which operates in a pulsed fashion and utilises magnetohydrodynamics (MHD) to accelerate high-temperature gases.

To provide the necessary propulsive energy, suitable sources of electrical energy, such as nuclear power generating plants (fission, controlled fusion), have to be carried along in rockets which rely on electrical propulsion.

	chemical thermal rockets				nuclear thermal rockets
kind of substance	solid	liquid	liquid-liquid	solid-liquid (lithergol)	nuclear energy + liquid
DIAGRAM OF ROCKET					
PROPELLANT SYSTEMS	double-base propellants compound propellants	CH_3NO_3 methyl nitrate H_2O_2 + potassium permanganate	propellant: kerosene, hydrazine, hydrogen oxygen reaction substances: O_2, F_2, HNO_3	low-oxygen propellant O_2HNO_3 H_2O_2	working fluid: hydrogen, ammonia, water
flow velocity	up to 3000 m/sec.	2000 m/sec.	up to 4500 m/sec.	2500 m/sec.	7000—30000 m/sec.

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